1, Mathematics — Todhunter, 1846
2, Physics — " , "
3, Astronomy — " , "

TD

AN

ELEMENTARY COURSE

OF

MATHEMATICS,

DESIGNED PRINCIPALLY FOR STUDENTS OF THE UNIVERSITY OF CAMBRIDGE.

BY THE

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PREFACE.

THE design of the following work will be understood by reference to the following extracts from a Grace of the Senate, passed in May, 1846, which will regulate the examination of Candidates for Mathematical Honours in January, 1848, and succeeding [illegible]

By the same Author, and to be published shortly by

MESSRS J. & J. J. DEIGHTON,

A Collection of Problems and Examples, adapted to the Elementary Course of Mathematics.

and Examiners, taking into account the Examination of all the eight days, shall arrange all the Candidates who have been declared to deserve Mathematical Honours into the three classes of Wranglers, Senior Optimes and Junior Optimes, as has been hitherto usual; and that these classes be published in the Senate House at nine o'clock on the Friday morning preceding the general B.A. Admission.

3. That the subjects of the Examination on the first three days shall be those contained in the following Schedule:—

EUCLID. Book I to VI. Book XI, Props. I to XXI. Book XII, Props. I, II.

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THE design of the following work will be understood by reference to the following extracts from a Grace of the Senate, passed in May, 1846, which will regulate the examination of Candidates for Mathematical Honours in January, 1848, and succeeding years.

It was determined by the Grace referred to,

1. That Questions and Problems being proposed to the Questionists on eight days, instead of six days as at present, the first three days be assigned to the more elementary, and the last five to the higher parts of Mathematics: that after the first three days, there shall be an interval of eight days; and that on the seventh of these days the Moderators and Examiners shall declare, what persons have so acquitted themselves as to deserve Mathematical Honours.

2. That those who are declared to have so acquitted themselves, and no others, be admitted to the Examination in the higher parts of Mathematics; and that after that Examination, the Moderators and Examiners, taking into account the Examination of all the eight days, shall arrange all the Candidates who have been declared to deserve Mathematical Honours into the three classes of Wranglers, Senior Optimes and Junior Optimes, as has been hitherto usual; and that these classes be published in the Senate House at nine o'clock on the Friday morning preceding the general B.A. Admission.

3. That the subjects of the Examination on the first three days shall be those contained in the following Schedule:—

EUCLID. Book I to VI. Book XI, Props. I to XXI. Book XII, Props. I, II.

ARITHMETIC and the elementary parts of ALGEBRA; namely, the Rules for the fundamental Operations upon Algebraical Symbols, with their proofs; the solution of simple and quadratic Equations; Arithmetical and Geometrical Progression, Permutations and Combinations, the Binomial Theorem, and the principles of Logarithms.

The elementary parts of PLANE TRIGONOMETRY, so far as to include the solution of triangles.

The elementary parts of CONIC SECTIONS, treated geometrically, together with the values of the Radius of Curvature, and of the Chords of Curvature passing through the Focus and Centre.

The elementary parts of STATICS, treated without the Differential Calculus; namely, the Composition and Resolution of Forces acting in one plane on a point, the Mechanical Powers, and the properties of the Centre of Gravity.

The elementary parts of DYNAMICS, treated without the Differential Calculus; namely, the Doctrine of Uniform and Uniformly Accelerated Motion, of Falling Bodies, Projectiles, Collision, and Cycloidal Oscillations.

The 1st, 2nd, and 3rd Sections of NEWTON'S PRINCIPIA; the Propositions to be proved in Newton's manner.

The elementary parts of HYDROSTATICS, treated without the Differential Calculus; namely, the pressure of non-elastic Fluids, specific Gravities, floating Bodies, the pressure of the Air, and the construction and use of the more simple Instruments and Machines.

The elementary parts of OPTICS, treated geometrically: namely, the laws of Reflection and Refraction of Rays at plane and spherical surfaces, not including Aberrations; the Eye; Telescopes.

The elementary parts of ASTRONOMY; so far as they are necessary for the explanation of the more simple phenomena, without calculation.

4. That in all these subjects, Examples, and Questions arising directly out of the Propositions, shall be introduced into the Examination, in addition to the Propositions themselves.

5. *(This article refers merely to the days and hours of Examination, and is therefore omitted.)*

6. That the Moderators and Examiners shall be authorized to declare Candidates, though they have not deserved Mathematical Honours, to have deserved to pass for an Ordinary Degree, so far as the Mathematical part of the Examination for such degree is concerned; and such persons shall accordingly be excused the Mathematical part of the Examination for an Ordinary Degree, and shall only be required to pass in the other subjects, namely, in the parts of the Examination assigned in the Schedule to the last two days: but such excuse shall be available to such persons only for the Examination then in progress.

When the preceding regulations had passed into law, it struck me very forcibly that, in order to carry out the expressed wishes of the University, it would be desirable, if not necessary, that a short course of mathematics should be compiled, of which the Schedule agreed upon by the Senate should be, as it were, the table of contents. It appeared to me that, with regard to several of the subjects, there were no books in use, which would put before the student the portions to which it would be necessary for him to devote his attention, without an accumulation of other matter which would be likely to confuse and perplex. I mentioned the necessity of such a book several times in the course of conversation, and found that others agreed as to the want, but I did not hear of any one who seemed disposed to undertake the labour requisite for its supply. Under these circumstances I determined to attempt the task myself, trusting that the intention would be appreciated, however much the execution of the design might fall short of my own hopes, or the requirements of the case.

For indeed it is a task of no ordinary degree of diffi-

culty, to write an elementary work upon an abstruse subject; points, which appear to the writer plain and intelligible without explanation, sometimes assume a very different aspect to the beginner, and difficulties of which the author is scarcely aware, may be of huge dimensions to a mind not already familiarized with the mode of thinking which belongs to each particular subject. Hence it has come to pass that so few elementary works have long retained their ground; and hence also, an author may conclude the expediency of endeavouring, so far as he may, to follow in the steps of those few who have shewn an aptitude for this kind of writing. Time has, I think, proved that, of all works which Cambridge has produced, that which the most nearly fulfils the conditions of a perfect elementary treatise, is *Wood's Algebra*, a work which it is impossible too much to admire for its simplicity and admirable perspicuity. In writing on Algebra, therefore, I have endeavoured, as far as possible, to take Dr Wood's treatise as a model: and, indeed, in all other parts of my work, where the nature of the case allowed, I have endeavoured, though I fear not always successfully, to keep the same example in view.

I may remark also, in general, that it appeared to me that the worst fault into which I could fall, in such a work as the present, would be an affectation of originality. Originality belongs to the progression of science, but not to the exposition of those portions which may be regarded as permanent. I have, therefore, endeavoured to deviate as little as possible from the methods pursued in those books which appeared to me, on the whole, to be the best and the most generally acknowledged.

In each treatise I have included all propositions which, according to my judgment, can fairly be included

in the intention of the Grace of the Senate, omitting however some which are usually given, but which are only applications of, or deductions from, the fundamental propositions. To take an example, in the treatise on Statics I have not given the investigations for pullies, when their own weight is taken into account, nor when the strings of the pullies are not parallel, because these are only deductions from the simplest case; and it would seem to be unadvisable to load a treatise, intended to be of the simplest description, with deductions and applications which may be indefinitely multiplied. Nevertheless my judgment may have led me into error; and, indeed, the only authorised comment on the meaning of the Grace, will be the Examination Papers of 1848 and the few succeeding years: the character of those papers may perhaps render it necessary to modify some portions of my work in a succeeding edition, should that be called for.

I will, however, be bold to give my opinion, that the success of the new scheme of examination depends, to a considerable extent, upon the Grace of the Senate being interpreted in the narrowest manner possible. The great number of the subjects of which a knowledge is required, renders it impossible for students of no great ability, or whose attention is principally devoted to other academical pursuits, to obtain a very extended knowledge of each; a very real and useful knowledge may doubtless be acquired, but it can scarcely be expected to be adequate to the task of answering questions proposed, unless those questions be almost confined to that small number of propositions which may be considered to be classical in each subject.

A few remarks may be made respecting certain of the following treatises, and the plan upon which they have been written.

The treatise on Algebra has, as I have before observed, been formed as much as possible, on the model of that by the late Dr Wood. I have given Euler's proof of the Binomial Theorem for fractional and negative indices, as being at once the most elegant and the most useful as a mental exercise.

In the treatise on the Conic Sections, I have principally followed the demonstrations given by Mr Hustler, partly because they appeared to me as elegant as could be desired, and partly because that work having been long published, and being usually counted the text book in this subject, I thought it well to deviate as little as possible from the beaten track. I have, however, abbreviated to some extent by omission, making it in general a necessary condition of the admission of any proposition that it should be necessary to the understanding of the first three sections of Newton's Principia; by this and other means, I have endeavoured to render as little formidable as might be a subject confessedly difficult and unpalatable.

In the subject of Mechanics, it is more difficult than in either of the preceding to determine precisely the limits to which questions may, according to the Grace of the Senate, extend. On principles to which I have before referred, I have made the treatises on both Statics and Dynamics as brief as possible.

In giving an English version of the first three sections of the *Principia*, I have endeavoured to adhere, as nearly as circumstances would allow, to the original, only giving the demonstrations a form more convenient for the purposes of the student; and any interpolations of my own have been enclosed in brackets. In one instance (Lemma VI.) I have, after the example of a version much used in the University, substituted a different mode of demonstration for the very short method given by Newton; but

the mode of demonstration adopted may be now, I think, considered classical, as having been given in the notes to the Jesuits' edition of the *Principia*. One proposition of the second section, which it has been usual in Cambridge to omit, I have also omitted.

The theory of Cycloidal Oscillations has been made an appendix to Newton's second section, because the time of oscillation can be deduced very simply from Prop. X., and this method at once avoids a complicated investigation, and also appears to exhibit clearly the nature of the dynamical action of gravity.

In writing on Optics, I have not construed the terms of the Grace, "treated geometrically," as intended to exclude all reference to the representation of lines by symbols with appropriate signs indicative of their direction; because, this being a method with which the student will have become familiar long before he reaches the science of Optics, and which in that science is as convenient as in any, it seemed to me quite certain that it could not be the design of the Grace to forbid the use of a convention so simple and useful. I have not, however, trusted to algebraical generalizations of formulæ proved in particular cases, but have either proved them for each case or pointed out how they may be proved, and have then explained how, by the convention respecting the meaning of the minus sign, all cases may be comprehended in one formula. I have, in fact, rather considered the term *geometrical* as exclusive of any new mode of investigation, such as that of co-ordinates, than as intended to expel from Optics conventions which are recognised in every treatise on Trigonometry, and which even Dr Wood has not wholly excluded. Dr Wood's treatise on Optics was carefully examined with reference to this

part of my work, but the great improvements in the mode of treating the subject, which have been made since the publication of that book, and of which I deemed it wrong to deprive the student, rendered it of comparatively little service.

The most difficult subject to treat in a clear and elementary manner is, I think, the last, Astronomy; yet is there none upon which a treatise adapted to the new scheme of examination was more required. For hitherto Astronomy has only been studied for the purposes of the Senate-house examinations by a few, and those the more advanced of the candidates for Honours; and consequently no book has been called forth containing those elementary portions of the subject, which will henceforth be required, in a form adapted to the needs of the majority of students. With regard to the manner in which I have treated the subject, I may remark, that I have not construed the words of the Grace, "without calculations," so as to exclude the introduction of mathematical symbols when by that means distinctness of explanation could be most easily attained; and also I may say that to no subject more than this does the remark apply which I have already made, namely, that it is difficult to determine what is and what is not intended to be included in the schedule of subjects, and that the difficulty will only disappear under the light of the examination papers of 1848 and a few succeeding years.

In conclusion, I will only add that the following pages have been written under a pressure of other engagements, and at a consequent degree of personal inconvenience, which would only have been submitted to under the conviction of the necessity of such a work in the present state of the University. I am well aware of the imper-

fection which must necessarily attach to the performance of my task, but which yet, I trust, will not be such as to affect to a material extent the utility of my book to those for whom it has been principally written.

H. GOODWIN.

DEC. 1846.

P.S.—I hope shortly to publish a set of Problems and Examples suitable to the plan of the present course.

H. G.

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29	6	$6x$	64.
33	19	3606	3600.
34	11	$\frac{3ab+b^3}{3a^2}$	$\frac{3ab^2+b^3}{3a^2}$.
51	14	$\frac{\sqrt{3}+5}{2}$	$\frac{3+\sqrt{5}}{2}$.
132	2	P	S.
134	last but one	measuring	reasoning.

146. Prop. IV. The following commencement of the demonstration is free from an objection which applies to that given in the text.

Because SYY' is a right angle; $\therefore$ ZCY' is a diameter of the auxiliary circle, and CZ, CY' are in the same straight line. Therefore SCY', CHZ are similar triangles, and $CS = CH$; therefore they are equal in all respects......&c.

A similar remark applies to Prop. IV. page 149, which will commence thus: 54

Because YZY' is a right angle; $\therefore$ YCY' is a diameter of the auxiliary circle, and CY, CY' are in the same straight line. Therefore SCY, HCY' are similar triangles, and $CS = CH$, therefore they are equal in all respects...,, &c.

PAGE	LINE	FOR	READ
156	21	ellipse has to the circle	circle has to the ellipse.
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209	1	$W_2 k_2$	$W_2(k_2 - A b_1)$.
220	14	required	acquired.
241	4 and 5 should stand thus:		

$$\text{Now } SPB = QPB - QPS = QPB - MPQ$$
$$= a - 90^\circ + a = 2a - 90.$$

PAGE	LINE	FOR	READ
	7	SP	$-SP$.
	9	$\frac{V^2}{2g}$	$-\frac{V^2}{2g}$.

(Note: the quantity $-\frac{V^2}{2g}\cos 2a$ will be *positive*, because in the figure a has been supposed greater than 45°).

PAGE	LINE	FOR	READ
248	12	$V' + \frac{R}{M}$	$V' + \frac{R}{M'}$.
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ALGEBRA.

ALGEBRA.

1. Algebra in its most comprehensive sense may be defined as being the science of reasoning by means of general written symbols.

In its simplest form we may consider algebra as a more general species of arithmetic, in which the reasoning and the operations refer to certain general representatives of numbers, instead of being applied to the symbols of specific numbers, viz. the digits 1, 2, 3...which are the subject of arithmetical reasoning. The science of algebra is of a far more general and comprehensive character than this; but even restricting our views to such a science as would be formed by a substitution of symbols of numbers in general for the nine digits, we may see at once the greatness of the advance which we have made beyond the limits of mere arithmetic, because any rule which has been established by means of algebraical symbols will be universally true, since the symbols may represent any numbers whatever; whereas it is difficult to establish a rule by means of operations which deal with particular numbers only. This remark will be understood better as the student proceeds.

In the following treatise it will be assumed that the student has already made himself acquainted with arithmetic; but some of the rules and operations of that science will be proved and explained as applications of algebra. Indeed, the theory of some of the arithmetical processes will seldom be seen distinctly, until the mind has been informed by the more general science.

2. The symbols used to denote *numbers* or *quantities* are usually the letters of the alphabet; and it is the common practice to express known or determined quantities by the early letters, as a, b, c..., and unknown by the latter, as x, y, z; but this rule is purely conventional, and need not be strictly followed.

It will be convenient here to enumerate and explain the various signs which are used in algebra.

3. + *Plus*, signifies that the quantity to which it is prefixed must be added. Thus $2 + 3$ is the same thing as 5; using letters, $a + b$ represents the *sum* of a and b, whatever are the values of a and b, or a and b *added together*.

4. − *Minus*, signifies that the quantity to which it is prefixed must be subtracted. Thus $3 - 2$ is the same thing as 1; and $a - b$ represents a with b taken from it.

5. Since the signs + and − prefixed to a quantity b indicate addition and subtraction, they would seem to imply some antecedent quantity to which b is to be added or from which it is to be subtracted; they are used however without this restriction, and for the present it will be sufficient for the student to consider $+ b$ as a quantity which *is to be* added, and $- b$ as a quantity which *is to be* subtracted. One of the signs + and − is supposed to be prefixed to every algebraical quantity: $+ a$ is termed a *positive* quantity, $- a$ a *negative* quantity. When $+ a$ is not preceded by another quantity, it is usual for shortness' sake to omit the +, and to write simply a.

A simple illustration of the meaning of a negative quantity may here be of service. A *debt* may be regarded as a negative quantity, inasmuch as it is a quantity *to be subtracted* in case of there being any property, which is *positive*, from which to subtract it.

6. × *Into*, signifies that the quantities between which it stands are to be multiplied together: thus 2×3 is equivalent to 6.

This sign is frequently omitted, or its place supplied by a point: thus $a \times b$, ab, $a.b$, are equivalent.

7. When the same quantity is multiplied into itself several times, the product is represented in an abbreviated form, by placing above the quantity a figure indicating the number of times that it is repeated: thus $a \times a$ is written a^2,

and $a \times a \times a$, a^3: a^2 is called the second *power*, or the square of a; a^3 the third *power*, or the cube; a^4 the fourth power, and so on: a^1 is the same thing as a.

The figure which indicates the *power* of a quantity is called its *index* or *exponent*. The meaning of the terms *power*, *index*, *exponent*, will be hereafter much extended. (See Art. 25.)

8. $\div$ *Divided by*, signifies that the former of two quantities between which it is placed is to be divided by the latter. Thus $6 \div 2$ is equivalent to 3.

Division is however more generally represented by writing the two quantities as a vulgar fraction: thus $a \div b$ is written $\frac{a}{b}$, and is commonly read for shortness' sake thus, *a by b*.

For another mode of representing division, see Art. 25.

9. The *difference* of two quantities is sometimes represented by the sign $\sim$: thus $a \sim b$ means $a - b$ or $b - a$ according as a is greater or less than b.

10. When several quantities are inclosed in a *bracket*, thus $(a + b - c)$, it is intended that any sign prefixed or affixed to the bracket should apply to all the quantities included by it: thus $a - (b + c)$ means that b and c are both to be subtracted from a, $(a + b)^2$ means that the sum of $a + b$ is to be multiplied by itself, $(2 + 1)(3 + 2)$ is equivalent to 3×5 or 15.

A *vinculum* or line drawn over several quantities, thus $\overline{a + b + c}$, is in all respects equivalent to a bracket.

11. $=$ *Equals*, signifies that the quantities between which it is placed are equal to each other: thus the symbolical sentence $2 + 3 = 5$ expresses the fact that 2 and 3 added together make 5.

12. An algebraical sentence expressing equality between two expressions, such as $x + 1 = 2$, or $x^2 + x + 1 = 0$, is called an *equation*.

13. $\therefore$ is an abbreviation for the word *therefore*.

14. The *square root* of a quantity is represented by writing over the quantity the sign $\sqrt[2]{\ }$ or more briefly $\sqrt{\ }$, which is in fact a corruption of the letter *r* standing for *radix* or *root*, and is termed a *radical:* thus $\sqrt[2]{a}$ or $\sqrt{a}$ means the square root of *a*. Similarly the *cube root* is denoted by $\sqrt[3]{a}$, the *fourth* root by $\sqrt[4]{a}$, and so on.

A more convenient mode of denoting these quantities will be found hereafter. (See Art. 25.)

A quantity under a *radical* sign, the root of which cannot be extracted, is called an *irrational* quantity or a *surd:* thus $\sqrt{3}$ is a surd quantity. Quantities which involve no surds are called *rational*.

15. The number or quantity by which any other quantity is multiplied, is frequently called its *coefficient:* thus in the quantities ax, $7y$, a and 7 may be called the coefficients of x and y respectively. When no coefficient is prefixed to a letter, 1 is always understood.

16. Any combination of symbols is called an algebraical *expression*, and sometimes an algebraical *formula*.

An expression is said to be of n dimensions with respect to any letter, when the highest power of that letter in the expression is n: thus $a^3 + 2a^2 + 1$ is of *three* dimensions with respect to a.

When an expression is composed of *two* quantities connected by the sign + or −, it is called a *binomial* expression; when of *three*, a *trinomial;* when of several, a *polynomial:* $a + b$ is a *binomial*, $a + b + c + d$ a *polynomial*.

When an expression is composed of quantities connected by the sign ×, (either expressed or understood,) the several quantities are called *factors* of the expression: thus a, b, c, are *factors* of abc.

17. *Similar* or *like* algebraical quantities are such as differ only in the value of their numerical coefficient: thus $2a$ and $5a$ are *like* quantities.

18. One quantity is said to be a *multiple* of another when the one can be divided by the other without remainder, or when the one contains the other as a *factor:* thus $4a$ is a multiple of a, and ax of x.

19. One quantity is said to be a *measure* of another, when the former will divide the latter without remainder.

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20. The student will find it advantageous, before proceeding further, to fix in his mind the knowledge acquired from the preceding articles, by practising upon some simple examples. And it may be worth while in this place to remark, that the science of algebra, as well as others which will be treated of hereafter, are most easily to be acquired by the practice of them; and indeed it may be said to be almost impossible to acquire and retain a perfect familiarity with algebraical theorems, except through the medium of a considerable amount of industry expended on the working out of examples. Wherefore, once for all, the student is earnestly requested after reading any new rule or theorem to turn to the examples illustrative of it, and work as many as possible before proceeding further.

ADDITION.

21. Rule. *The addition of algebraical quantities is performed by connecting those that are unlike with their proper signs, and collecting those that are like into one sum.*

The addition of algebraical quantities would in fact be performed by writing them one after another, and connecting them by the sign $+$; but the preceding rule indicates the mode of reducing the sum so written down to its simplest form.

When *like* quantities occur with different signs, their algebraical sum is found by taking the smaller coefficient from the greater and prefixing the sign of the greater: thus $4a-3a$ would be written a, and $-4a+3a$ would be written $-a$.

It will be seen that addition thus considered is a very different operation, in some respects, from arithmetical addition,

since in arithmetic to *add* is always to *increase*, but in algebra to *add* is only to *connect a series of quantities with their proper signs;* and a quantity may therefore be decreased by having another, which is negative, added to it.

The order in which the result of the addition of several quantities is written down is immaterial, but it is usual *cæteris paribus* to follow the order of the alphabet: thus we should write $a + b + c$, not $a + c + b$.

The following are examples of addition, which the student may verify:

$$\begin{array}{rr} & 2a + b \\ & a + 3b \\ \hline \text{Sum} & 3a + 4b \\ \hline \end{array} \qquad \begin{array}{r} 2a - b \\ a + 3b \\ \hline 3a + 2b \\ \hline \end{array}$$

$$\begin{array}{r} -2a + b \\ a + 3b \\ \hline -a + 4b \\ \hline \end{array} \qquad \begin{array}{r} 2a - b \\ a - 3b \\ \hline 3a - 4b \\ \hline \end{array}$$

$$\begin{array}{r} a^2 + 2ab + 3b^2 \\ -2a^2 - ab + 4b^2 \\ 3a^2 + 3ab - 2b^2 \\ \hline 2a^2 + 4ab + 5b^2 \\ \hline \end{array} \qquad \begin{array}{r} -a + b + c + d \\ a - b + c + d \\ a + b - c + d \\ a + b + c - d \\ \hline 2a + 2b + 2c + 2d \\ \hline \end{array}$$

$$\begin{array}{l} a^2 + bc + cd \\ b^2 + bd + c^2 \\ a^2 + b^2 + c^2 \\ \hline 2a^2 + 2b^2 + 2c^2 + bc + bd + cd \\ \hline \end{array}$$

$$\begin{array}{l} ax + by + cz \\ bx - cy - az \\ -cx + ay - bz \\ \hline (a + b - c)x + (b - c + a)y + (c - a - b)z \\ \hline \end{array}$$

SUBTRACTION.

22. RULE. *One algebraical quantity may be subtracted from another by changing its sign and adding it to the other.*

The reason of this rule is apparent; for to subtract $+a$ is by definition the same thing as to add $-a$; and to subtract a quantity which itself ought to be subtracted is nothing else than to add that quantity. Using an illustration already adopted, we may say that to *subtract* a *debt* is the same thing as to *add* to *property*.

Hence also we see that $-$ prefixed to a bracket changes the sign of all the quantities in that bracket, so that $-(a+b-c)$ is equivalent to $-a-b+c$.

EXAMPLES.

$$\begin{array}{r} 2a+b \\ a+3b \\ \hline a-2b \end{array} \qquad \begin{array}{r} 2a-b \\ -a+3b \\ \hline 3a-4b \end{array} \qquad \begin{array}{r} a+b-2c \\ a-b+c \\ \hline 2b-3c \end{array}$$

Difference (first row): $a-2b$, $3a-4b$, $2b-3c$

$$\begin{array}{r} 2a^2+3ab-b^2 \\ a^2-4ab+3b^2 \\ \hline a^2+7ab-4b^2 \end{array} \qquad \begin{array}{r} a+b-(c+d) \\ a-(b-c)+d \\ \hline 2b-2c-2d \end{array}$$

$$\begin{array}{r} (a+b)x^2-(c-d)y^2 \\ (b+c)x^2+(a+d)y^2 \\ \hline (a-c)x^2-(a+c)y^2 \end{array}$$

MULTIPLICATION.

23. *Unlike* quantities cannot be multiplied together any further than by connecting them by the sign $\times$ or writing them together without sign. Thus a multiplied by b gives

ab or *ba*, for it is indifferent whether we consider *a* to be multiplied by *b* or *b* by *a*.

But *like* quantities are multiplied together by ***adding their indices***. Thus $a^3 \times a^4 = a^7$, for a^3 signifies aaa and a^4 signifies $aaaa$, therefore $a^3 \times a^4 = aaa \times aaaa = aaaaaaa = a^7$ by definition, for a^7 means nothing else but a multiplied into itself *seven* times.

The *sign* of the product of two quantities is determined by this rule; viz. the product of two quantities affected by the *same* sign is *positive*, of two affected by *different* signs *negative*.

$$\text{Thus } +a \times +b = +ab;$$
$$-a \times +b = -ab;$$
$$+a \times -b = -ab;$$
$$-a \times -b = +ab.$$

This rule may be thus explained:

(1) $+a \times +b$ signifies that a is to be added b times, which is the same as adding 1 ab times, therefore the result is $+ab$.

(2) $+a \times -b$ signifies that b is to be subtracted a times, which is the same thing as subtracting 1 ab times, therefore the result is $-ab$.

(3) $-a \times +b$ signifies that a is to be subtracted b times, therefore, as in the last case, the result is $-ab$.

(4) $-a \times -b$ may be interpreted to mean that $-a$ is to be subtracted b times, or that $-ab$ is to be *subtracted*, or that ab is to be *added*, therefore the result is $+ab$.

Numbers are multiplied together as in common arithmetic: thus $2a \times 3b$ is not written $2 \times 3ab$ but $6ab$.

What has been said hitherto applies chiefly to the multiplication of *simple* algebraical quantities, or expressions of *one* term only. When two polynomials are multiplied together, each term in the multiplicand must be multiplied by each term in the multiplier, and the sum of all such products (arranged as is most convenient) will be the complete product required.

It is usual in algebraical multiplication to commence with the term on the left hand of an expression, instead of commencing on the right as in arithmetic.

When the same letter occurs in an expression with different indices it is usual, and in most cases of the application of algebra necessary, to arrange the expressions according to the powers of that letter: thus the expression $1-3x+3x^2-x^3$ ought not to be written $1+3x^2-3x-x^3$, but it may with propriety be written as we have given it, in which case it is said to be arranged according to *ascending powers of x*; or it may be written thus $-x^3+3x^2-3x+1$, in which case it is said to be arranged according to *descending powers of x*. The student cannot be too careful in attending to the proper *arrangement* of expressions.

EXAMPLES.

$$\begin{array}{lllll} & a+b & & & \\ & c+d & & & \\ \hline & ac+bc & & & \\ & & ad+bd & & \\ \hline \text{Product} & ac+bc & +ad+bd & & \\ \hline \end{array}$$

$$\begin{array}{ll} a+b & \\ a-b & \\ \hline a^2+ab & \\ & -ab-b^2 \\ \hline a^2 & \quad\quad -b^2 \\ \hline \end{array}$$

$$\begin{array}{llll} a+ & bx+ & cx^2 & \\ a- & bx+ & cx^2 & \\ \hline a^2+ & abx+ & acx^2 & \\ & -abx- & b^2x^2-bcx^3 & \\ & & acx^2+bcx^3+c^2x^4 & \\ \hline a^2 & & +(2ac-b^2)x^2+c^2x^4 & \\ \hline \end{array}$$

In this example the product has been arranged according to *ascending powers of x*, because the multiplicand and multiplier were so arranged; and on this account the two terms involving x^2, viz. $2acx^2$ and $-b^2x^2$, have been collected into

one term, and the combined quantity $2ac - b^2$ is considered as the coefficient of x^2.

$$\begin{array}{l}
x^2 + 2x - 1 \\
x^2 - 2x + 1 \\
\hline
x^4 + 2x^3 - x^2 \\
\quad - 2x^3 - 4x^2 + 2x \\
\qquad\qquad x^2 + 2x - 1 \\
\hline
x^4 \qquad - 4x^2 + 4x - 1 \\
\hline
\end{array}$$

DIVISION.

24. Division being the inverse of multiplication, its rules may be deduced from those of multiplication.

Unlike quantities can be divided one by another only by writing one under the other in the form of a vulgar fraction.

But *like* quantities can be divided one by another by subtracting the index of the divisor from that of the dividend. Thus $a^3 \div a = a^2$, because as we have seen $a \times a^2 = a^3$.

The rule of signs is this; the division of quantities of *like* signs gives a *positive* quantity, and of *unlike* signs a negative.

Sometimes the division of unlike quantities can be partially effected: thus $\frac{a^2b^2c}{abd} = \frac{abc}{d}$, where the dividend can be divided by the factors a and b of abd, but not by the factor d.

When a polynomial is to be divided by a simple quantity each term of the polynomial must be divided by it, and the sum of the terms so found affected with their proper signs will be the quotient.

The process of dividing one polynomial by another is one of greater difficulty, but is rendered sufficiently simple by its analogy to long division in common arithmetic. The

first step is to arrange the divisor and dividend according to either ascending or descending powers of some letter common to the two; the division of the first term of the dividend by the first term of the divisor gives the first term of the quotient; multiply the divisor by this term, and subtract the product from the dividend; bring down as many more of the terms of the dividend as may be required, and repeat the process until all the terms have been brought down.

The only point in this rule which seems to require explanation, is the arranging of the expressions according to powers of some common letter. The reason may be given thus: division is the inverse of multiplication, and in order to make division successful we must be sure that we follow exactly the reverse steps of some particular mode of multiplying, for two expressions may be multiplied together in many different ways according to the arrangement which we choose to adopt; now we are sure of following an exactly reverse process by attending to the rule of arrangement which has been given, for the quotient and divisor may be conceived to have been multiplied together according to this rule to form the dividend.

EXAMPLES.

$$\begin{array}{rl}
a+b)\ a^2 - b^2 & (a - b \quad \text{Quotient} \\
\underline{a^2 + ab} & \\
-ab - b^2 & \\
\underline{-ab - b^2} &
\end{array}$$

$$\begin{array}{rl}
x-y)\ x^3 - 3x^2y + 3xy^2 - y^3 & (x^2 - 2xy + y^2 \\
\underline{x^3 - x^2y} & \\
-2x^2y + 3xy^2 & \\
\underline{-2x^2y + 2xy^2} & \\
xy^2 - y^3 & \\
\underline{xy^2 - y^3} &
\end{array}$$

$$
\begin{array}{l}
x + y)\ x^n + y^n \quad (x^{n-1} - x^{n-2}y + x^{n-3}y^2 - \&c. \\
\quad \underline{x^n + x^{n-1}y} \\
\qquad - x^{n-1}y + y^n \\
\qquad \underline{- x^{n-1}y - x^{n-2}y^2} \\
\qquad\qquad\qquad x^{n-2}y^2 + y^n \\
\qquad\qquad\qquad \underline{x^{n-2}y^2 + x^{n-3}y^3} \\
\qquad\qquad\qquad\qquad - x^{n-3}y^3 + y^n \\
\qquad\qquad\qquad\qquad\quad \&c. \quad \&c.
\end{array}
$$

The following example is given to shew the importance of *arrangement:*

$$
\begin{array}{l}
a + 1)\ a^2 + 2a + 1 \ (a + 1 \\
\qquad\quad \underline{a^2 + \ \ a} \\
\qquad\qquad\quad a + 1 \\
\qquad\qquad\quad \underline{a + 1}
\end{array}
$$

thus the operation terminates; but suppose we had proceeded thus:

$$
\begin{array}{l}
a + 1)\ 1 + 2a + a^2 \ \left(\dfrac{1}{a} - \dfrac{1}{a^2} + \&c.\right. \\
\qquad\quad \underline{1 + \dfrac{1}{a}} \\
\qquad\quad - \dfrac{1}{a} + 2a \\
\qquad\quad \underline{- \dfrac{1}{a} - \dfrac{1}{a^2}} \\
\qquad\qquad\quad \dfrac{1}{a^2}\ \&c.
\end{array}
$$

and the operation would never come to an end.

ON THE MEANING OF NEGATIVE AND FRACTIONAL INDICES.

25. Hitherto a^n has been understood to signify $a \times a \times \ldots$ n times, and therefore n has been supposed to be a whole number. But it becomes a question whether it may not be possible to assign a meaning to the symbol a^n in other cases, whether for instance we may not assign a meaning to $a^{\frac{1}{2}}$, and to a^{-2}. In doing so the only thing to be attended to is, that no supposition be made contradictory to any thing which we have at present laid down, and it will manifestly be most convenient that the rules for multiplying and dividing such quantities as we speak of should be the same as in the case of a positive index. Suppose then we make this convention, that the rule of indices which has been proved in the case of positive integer indices (Art. 23) shall hold true in all cases; that is, let us *assume* that *universally*

$$a^m \times a^n = a^{m+n} \qquad (1)$$

$$\text{and} \qquad \frac{a^m}{a^n} = a^{m-n} \qquad (2),$$

then it will be found that these assumptions, which contradict nothing preceding them, will be sufficient to determine the meaning of a^n when n is fractional or negative. For we have by (1)

$$a^{\frac{1}{2}} \times a^{\frac{1}{2}} = a^{\frac{1}{2}+\frac{1}{2}} = a,$$

$$\text{but } a^{\frac{1}{2}} \times a^{\frac{1}{2}} = (a^{\frac{1}{2}})^2;$$

$$\therefore \quad (a^{\frac{1}{2}})^2 = a,$$

$$\text{or } a^{\frac{1}{2}} = \sqrt{a}.$$

Or more generally,

$$a^{\frac{1}{p}} \times a^{\frac{1}{p}} \times \ldots\ldots p \text{ times} = a^{\frac{1}{p}+\frac{1}{p}\ldots\ldots \text{to } p \text{ terms}} = a;$$

but $a^{\frac{1}{p}} \times a^{\frac{1}{p}} \times \ldots\ldots p$ times $= \left(a^{\frac{1}{p}}\right)^p$;

$$\therefore \left(a^{\frac{1}{p}}\right)^p = a,$$

$$\text{or} \quad a^{\frac{1}{p}} = \sqrt[p]{a}.$$

Hence the symbol $a^{\frac{1}{p}}$ represents the p^{th} root of a, and $a^{\frac{q}{p}}$ represents the p^{th} root of a^q: thus $4^{\frac{1}{2}} = 2$, and $4^{\frac{3}{2}} = 8$. It will be seen also that $\sqrt{\sqrt{a}} = \sqrt{a^{\frac{1}{2}}} = a^{\frac{1}{2}\times\frac{1}{2}} = a^{\frac{1}{2^2}}$, and so on.

Again, suppose that in (1) we write $-n$ for n, then we have

$$a^m \times a^{-n} = a^{m-n};$$

but by (2)
$$\frac{a^m}{a^n} = a^{m-n};$$

$$\therefore a^m \times a^{-n} = \frac{a^m}{a^n},$$

which proves, that to multiply by a^{-n} is the same thing as to divide by a^n, or that $a^{-n} = \frac{1}{a^n}$; thus $2^{-\frac{1}{2}} = \frac{1}{\sqrt{2}}$, and $4^{-\frac{1}{2}} = \frac{1}{2}$.

Thus we have assigned to negative and fractional indices a meaning not inconsistent with any thing which precedes, and which will be found of great service. We shall therefore henceforth use the symbols $\sqrt[p]{a}$, $\frac{1}{a^n}$, and $a^{\frac{1}{p}}$, a^{-n} indifferently.

26. A rather remarkable consequence follows from (2), which will require a few words. If we suppose m and n to be equal, we have

$$\frac{a^m}{a^m} = a^{m-m},$$

$$\text{or} \quad 1 = a^0.$$

This result is, at first sight, somewhat paradoxical, nevertheless it must be received as true, being a legitimate deduction from

our previous assumptions; moreover, it is not wholly incapable of being interpreted in such a way as to make it intelligible. For suppose, *first*, that a is a number greater than 1; then if we extract its square root, it is evident that the square root will be less than the number itself; let the square root be again extracted and the number will be still further decreased; let this process be repeated a great many times, say a thousand, then the number will have decreased at each process, but will never be made less than 1, because if so, conversely a quantity less than 1 might be made greater than 1 by squaring, which is absurd: hence $a^{\frac{1}{2^{1000}}}$, which represents the square root of a taken a thousand times, must be very nearly $= 1$ but not quite. Again, *secondly*, let a be a number less than 1, and let the same process be performed upon it, then it is clear from the same kind of reasoning that the number will increase at each process, but that it can never become quite $= 1$: hence also in this case $a^{\frac{1}{2^{1000}}}$ nearly $= 1$ but not quite. But $\frac{1}{2^{1000}}$ is a very small quantity indeed, and it therefore appears that, whether a be greater or less than 1, when n is *very small*, a^n *very nearly* $= 1$, and hence we can see the meaning of the equation $a^0 = 1$.

THE GREATEST COMMON MEASURE.

27. Def. *The greatest common measure of two numbers is the greatest number which will divide both without remainder.*

The greatest common measure of two algebraical expressions must be defined rather differently.

Def. *Let two algebraical expressions be arranged according to descending powers of some common letter, then the factor of the highest dimension with respect to that letter, which divides both without remainder, is the greatest common measure of the two expressions.*

It would be more correct to speak of the *highest common divisor*, since the terms *greater* or *less* are not applicable to algebraical expressions, which are great or small according to the numerical values which we choose to assign to the letters involved: but, in accordance with established usage, the name of *greatest common measure* will be used.

28. *To investigate a rule for finding the greatest common measure of two algebraical expressions.*

Let A and B be two expressions arranged according to descending powers of some common letter, and let the highest power of that letter in B be not higher than the highest in A. Divide A by B, make the remainder the divisor and B the dividend, and so on, until you come to a quantity which will divide without remainder; this last divisor will be the greatest common measure required.

The operation indicated may be represented as under,

$$\begin{array}{llll}
B)\ A\ (p & & & \\
\ \ \ \underline{pB} & & & \\
\ \ \ C)\ B\ (q & & & \\
\ \ \ \ \ \ \ \underline{qC} & & & \\
\ \ \ \ \ \ \ D)\ C\ (r & & & \\
\ \ \ \ \ \ \ \ \ \ \ \underline{rD} & & & \\
\ \ \ \ \ \ \ \ \ \ \ \ \ \underline{0} & & &
\end{array}$$

from which we have the following relations,

$$A = pB + C \qquad (1),$$
$$B = qC + D \qquad (2),$$
$$C = rD \qquad (3).$$

Now it is manifest that any quantity (P) which measures two others, Q, R, will measure a quantity such as $mQ \pm nR$*,

* The expression $mQ \pm nR$ stands for $mQ + nR$ and $mQ - nR$: it is read thus, mQ *plus or minus* nR.

since it will divide each of the terms mQ and nR; but from (3) we see that D measures C, therefore it measures $qC + D$, that is, it measures B, by (2); therefore it measures $pB + C$, that is, it measures A, by (1). Hence D is a common measure of A and B. It is also the *greatest* common measure; for if not, let it be D'; then since D' measures both A and B it measures $A - pB$, that is, it measures C, by (1); therefore it measures $B - qC$, that is, it measures D, by (2): but D' cannot measure D if it be a quantity of higher dimensions, therefore D' is not a greater common measure than D, that is, D is the greatest common measure.

In order to render the preceding process successful, it will be necessary to modify the remainders, and also the given expressions, in such a manner as to avoid fractional coefficients in all the terms which occur. We may do this by multiplying by any quantity which does not introduce a new common measure, since it is clear that the proof which has been given, will not be affected by supposing the expressions so modified. Also, any factor which is found to belong to one remainder, and not to the other which is used as the divisor or dividend to it, should be omitted.

Ex. Find the greatest common measure of

$$3x^4 + 2x^3 + 3x + 2 \text{ and } 4x^3 + 10x^2 + 4x - 2.$$

The first thing to be done is to reject the factor 2 which belongs to the latter quantity and not to the former; then the operation is continued thus:

$$
\begin{array}{rl}
 & 3x^4 + 2x^3 + 3x + 2 \\
 & 2 \\
2x(2x^3 + 5x^2 + 2x - 1) & 6x^4 + 4x^3 + 6x + 4 \;(3x \\
= 4x^3 + 10x^2 + 4x - 2 & 6x^4 + 15x^3 + 6x^2 - 3x \\
 & -11x^3 - 6x^2 + 9x + 4 \\
 & 2 \\
 & -22x^3 - 12x^2 + 18x + 8 \;(-11 \\
x^3 + 5x^2 + 2x - 1 \,)\, -11 \,)\, = & -22x^3 - 55x^2 - 22x + 11 \\
 & 43x^2 + 40x - 3
\end{array}
$$

$$
\begin{array}{rrrrrl}
 & 2x^3 + & 5x^2 + & 2x - & 1 & \\
 & 43 & & & & \\
\hline
43x^2 + 40x - 3\,) & 86x^3 + & 215x^2 + & 86x - & 43 & (2x \\
 & 86x^3 + & 80x^2 - & 6x & & \\
\hline
 & & 135x^2 + & 92x - & 43 & \\
 & & 43 & & & \\
\hline
 & & 5805x^2 + & 3956x - & 1849 & (135 \\
 & & 5805x^2 + & 5400x - & 405 & \\
\hline
 & & & -1444x - & 1444 & \\
\hline
\end{array}
$$

(Rejecting the factor -1444)

$$
\begin{array}{rrrl}
x + 1\,) & 43x^2 + 40x & - 3 & (43x - 3 \\
 & 43x^2 + 43x & & \\
\hline
 & -3x & - 3 & \\
 & -3x & - 3 & \\
\hline
\end{array}
$$

Hence $x + 1$ is the greatest common measure required.

The example here given is one of considerable complication, but is worthy of attention as illustrating the peculiar difficulties besetting the search for the greatest common measure. The student is particularly advised to obtain facility in working examples under this rule before proceeding further.

29. The greatest common measure of three quantities A, B and C, is found thus; find D the greatest common measure of A and B, then the greatest common measure of D and C will be the greatest common measure required.

For every measure of A and B measures D, and therefore every measure of A, B and C measures C and D, and hence the highest measure of A, B and C will be the highest measure of C and D.

30. One application of the rule for finding the greatest common measure of two quantities is to the simplification of

fractions; they may frequently be simplified by inspection, but, if this cannot be done, find the greatest common measure of the numerator and denominator by the preceding method and divide both numerator and denominator by it.

THE LEAST COMMON MULTIPLE.

31. Def. *If* A *and* B *are two algebraic expressions arranged according to descending powers of some common letter, and* M *the quantity of lowest dimensions with respect to that letter which is divisible by both expressions, then* M *is the least common multiple of* A *and* B.

To find M, let m be any multiple of A and B, so that

$$m = pA = qB,$$

then by definition M will be that value of m for which p and q are of the *lowest* dimensions. But since $pA = qB$, we have $\frac{p}{q} = \frac{B}{A}$, and therefore the proper values of $\frac{p}{q}$ will be found by reducing $\frac{B}{A}$ to its *lowest* terms. Hence if D be the greatest common measure of A and B,

$$p = \frac{B}{D}, \qquad q = \frac{A}{D},$$

$$\text{and therefore } M = pA = qB = \frac{AB}{D}.$$

Hence we have this rule: Multiply the two expressions together and divide the product by the greatest common measure, the quotient will be the least common multiple.

In practice it is better to divide one of the quantities by the greatest common measure, and multiply the other by the quotient.

Example. Find the least common multiple of $x^3 - 2x + 1$ and $x^4 - x^3 - x + 1$.

$$
\begin{array}{r|l}
x^3 - 2x + 1) & x^4 - x^3 - x + 1 \;(x - 1 \\
 & x^4 - 2x^2 + x \\ \hline
 & -x^3 + 2x^2 - 2x + 1 \\
 & -x^3 + 2x - 1 \\ \hline
 & 2x^2 - 4x + 2
\end{array}
$$

$$
\begin{array}{r|l}
x^2 - 2x + 1) & x^3 - 2x + 1 \;(x + 2 \\
 & x^3 - 2x^2 + x \\ \hline
 & 2x^2 - 3x + 1 \\
 & 2x^2 - 4x + 2 \\ \hline
\end{array}
$$

$$
\begin{array}{r|l}
x - 1) & x^2 - 2x + 1 \;(x - 1 \\
 & x^2 - x \\ \hline
 & -x + 1 \\
 & -x + 1 \\ \hline
\end{array}
$$

Hence $x - 1$ is the greatest common measure.

Again

$$
\begin{array}{r|l}
x - 1) & x^4 - x^3 - x + 1 \;(x^3 - 1 \\
 & x^4 - x^3 \\ \hline
 & -x + 1 \\
 & -x + 1 \\ \hline
\end{array}
$$

$\therefore$ $(x^3 - 1)(x^3 - 2x + 1) = x^6 - 2x^4 + 2x - 1$ is the least common multiple required.

32. The least common multiple of three quantities A, B and C is found by determining the least common multiple D of any two of them, as A and B, and then finding the least common multiple of D and C.

ON FRACTIONS.

33. In arithmetic we define a fraction thus: A fraction is any part or parts of a unit or whole, and it consists of two members, a denominator and a numerator, whereof the former shews into how many parts the unit is divided, the latter shews how many of them are taken in the given case.

Thus $\frac{3}{5}$ denotes that the unit is divided into 5 parts and that 3 of them are taken; and more generally $\frac{a}{b}$ denotes that the unit is divided into b parts and that a of them are taken.

In algebra we cannot give exactly the same definition, for we call any quantity of the form $\frac{a}{b}$ a *fraction*, although a and b are not necessarily representatives of whole numbers, as they must be if the fraction be an *arithmetical* or *vulgar* fraction. What is meant by the algebraical fraction $\frac{a}{b}$ is simply this, that any quantity affected by it is to be *multiplied by* a, and *divided by* b; this definition however includes that given above for vulgar fractions, as it ought, because algebra is a more general science than arithmetic, and includes arithmetic in its rules.

34. *To add two or more fractions together, bring them to a common denominator, add the numerators for a new numerator, and take the common denominator for the new denominator.*

Let $\frac{a}{b}, \frac{c}{d}$ be two fractions,

$$\text{then } \frac{a}{b} + \frac{c}{d} = \frac{ad}{bd} + \frac{bc}{bd} = \frac{ad + bc}{bd};$$

for in the first place $\frac{a}{b} = \frac{ad}{bd}$, since to multiply and divide by

d cannot affect the value of the fraction; and in the next place $\frac{ad}{bd} + \frac{bc}{bd}$ indicates that a quantity is to be divided by bd and multiplied first by ad and then by bc, and that these two products are to be added together, whereas the expression $\frac{ad + bc}{bd}$ indicates division by bd and multiplication by $ad + bc$, but this is manifestly the same operation as in the former case, hence $\frac{ad}{bd} + \frac{bc}{bd} = \frac{ad + bc}{bd}$.

When any number of fractions are to be reduced to a common denominator, the rule (which requires no proof) is this: Multiply each numerator by every denominator except its own, and all the denominators together for a new denominator. This rule may be sometimes usefully superseded by the following: Find the least common multiple of the denominators, take this as the new denominator and its product by the several fractions for the several new numerators.

35. The rule for subtraction follows at once from that for addition, and is expressed by the formula

$$\frac{a}{b} - \frac{c}{d} = \frac{ad - bc}{bd}.$$

36. *To multiply two fractions together, multiply the numerators together for a new numerator and the denominators together for a new denominator.*

Let $\frac{a}{b}, \frac{c}{d}$ be the two fractions, then will $\frac{a}{b} \times \frac{c}{d} = \frac{ac}{bd}$. This is a consequence of the meaning of the symbols, for $\frac{a}{b} \times \frac{c}{d}$ signifies that the quantity c is to be divided by d, then multiplied by a, and then divided by b; and $\frac{ac}{bd}$ signifies that a and c are to be multiplied together and then divided by the product bd; hence the operations indicated in the two cases

are the same, the *order* of them is the only difference, but the order of operations has no effect on the result;

$$\therefore \frac{a}{b} \times \frac{c}{d} = \frac{ac}{bd}.$$

The same reasoning will shew that $a \times \frac{c}{d} = \frac{ac}{d}$.

It will be observed here as in other cases that a rather extended sense is given to the term *multiplication;* the student should understand by the term an algebraical operation of which multiplication in arithmetic is the *type.*

37. *To divide one fraction by another, invert the divisor and proceed as in multiplication.*

Since to *divide* is the inverse of to *multiply,* it follows that to divide by $\frac{a}{b}$ is the same thing as to multiply by $\frac{b}{a}$; and hence the rule.

ON THE THEORY OF DECIMAL FRACTIONS.

38. The principles of algebra which we have been developing are sufficient to enable us to explain and prove the rules of decimal fractions; and as we have been speaking of fractions, this will be a convenient place for introducing the subject.

39. DEF. *A decimal fraction is one which has* 10 *or some power of* 10 *for its denominator.*

Hence a decimal fraction will be represented algebraically by $\frac{N}{10^n}$; where n is the number of decimal places, and N is the whole number which the decimal would represent if we omitted the decimal point. Thus $1.37 = \frac{137}{10^2}$, in which case $N = 137$ and $n = 2$.

Having obtained this general symbolical representation of a decimal, all the rules will follow with great simplicity.

40. *To prove the rule for the multiplication of decimals.*

Let $\frac{M}{10^m}$, $\frac{N}{10^n}$ be two decimals; then

$$\frac{M}{10^m} \times \frac{N}{10^n} = \frac{MN}{10^{m+n}}.$$

The numerator of this last fraction shews that the decimal quantities are to be multiplied together as if they were whole numbers, the denominator that there must be $m + n$ decimal places, that is, as many as in the multiplier and multiplicand together.

The rule for division may be proved in like manner.

41. *To prove the rule for converting a vulgar fraction into a decimal.*

Let $\frac{A}{B}$ be the vulgar fraction; then

$$\frac{A}{B} = \frac{A \times 10^n \div B}{10^n} \text{ identically.}$$

The numerator shews that we are to add cyphers at pleasure to the right of A, (for that is the same thing as multiplying by 10^n,) and that we are then to divide by B; while the denominator indicates that as many places are to be marked off for decimals as we have added cyphers.

42. *To prove that every vulgar fraction must produce either a terminating or a recurring decimal.*

In the division of $A10^n$ by B every remainder must be less than B, therefore there can be at most only $B - 1$ different remainders; hence if no remainder becomes zero, that is, if the operation does not terminate, a remainder must recur within $B - 1$ operations at furthest; the figures in the quotient will then recur, and the result will be a recurring decimal.

43. *To determine the form of those vulgar fractions which produce terminating decimals.*

We have as before

$$\frac{A}{B} = \frac{A\,10^n \div B}{10^n}.$$

Now in order that the above may be a terminating decimal, B must divide $A\,10^n$ without remainder; but B cannot divide A since we suppose the fraction in its lowest terms, therefore it must divide 10^n. But it is easy to see that the only numbers which will divide 10^n are those which are made up of the factors 2 and 5, because these are the only two numbers which will divide 10. Hence B must have no other factors than 2 and 5, or speaking algebraically B must be of the form $2^p\,5^q$. For example, $\frac{5}{8}$ will produce a terminating decimal, because $8 = 2^3$; but $\frac{5}{24}$ will not, because 24 is divisible by 3.

These are the most simple propositions relating to decimals; others will be found in Arts. (60) and (107).

ON INVOLUTION AND EVOLUTION.

44. *Involution* is the process of multiplying a quantity by itself any number of times; *Evolution* is the finding of a quantity which being multiplied into itself a given number of times shall become equal to a given quantity, in other words evolution is the extraction of any *root* of a quantity.

45. The involution of a simple quantity is effected, as we have seen (Art. 25), by multiplying its index, and the evolution by dividing the index; for

$$(a^p)^q = a^{pq} \text{ and } \sqrt[q]{a^p} = a^{\frac{p}{q}}.$$

46. The involution of a polynomial is a very simple though frequently a laborious process, being a process of actual multiplication.

Ex. Find the square of $a + b$.

$$\begin{array}{l} a + b \\ a + b \\ \hline a^2 + ab \\ \quad\quad ab + b^2 \\ \hline a^2 + 2ab + b^2 = (a+b)^2 \\ \hline \end{array}$$

Find the cube of $a + b$.

$$\begin{array}{l} a^2 + 2ab + b^2 \\ a + b \\ \hline a^3 + 2a^2b + ab^2 \\ \quad\quad a^2b + 2ab^2 + b^3 \\ \hline a^3 + 3a^2b + 3ab^2 + b^3 = (a+b)^3 \\ \hline \end{array}$$

It will be easily seen that the labour increases very rapidly with the number of terms in the polynomial and also with the degree of the power.

47. The squaring of a polynomial is rendered very easy by the following theorem:

The square of any polynomial = the sum of the squares of the terms + twice the product of each two terms.

To prove this we observe that in the expression for $(a+b)^2$ the rule obviously holds, for $(a+b)^2 = a^2 + b^2 + 2ab$; now *suppose* the rule to hold for the polynomial $a + b + c + \ldots + l$, so that

$$(a + b + c + \ldots + l)^2 = a^2 + b^2 + c^2 + \ldots + l^2 + 2ab + 2ac + \ldots \quad (1)$$

then introducing another term m we must have

$$(a + b + c + \ldots + l + m)^2 = (\overline{a + b + c + \ldots + l} + m)^2$$
$$= (a + b + c + \ldots l)^2 + m^2 + 2(a + b + c + \ldots + l)\,m$$

(because $a+b+c+\ldots+l$ may all be considered as one quantity;)

$$= a^2 + b^2 + c^2 + \ldots + l^2 + 2ab + 2ac + \ldots\ldots\ldots\ldots$$
$$+ m^2 + 2am + 2bm + \ldots \text{ by (1)}$$

= the sum of the squares of the terms + twice the product of each two.

Hence *if* the theorem be true for a polynomial of any number of terms, it will be true for a polynomial of that number of terms increased by one: but the theorem *is* true for an expression of *two* terms, ∴ it is true for one of *three*, ∴ for one of *four*, ∴ &c., ∴ for any polynomial.

The mode of reasoning adopted in the preceding proposition is one which we shall have to make use of again, and is therefore worthy of attentive consideration.

Examples of the application of this theorem:

$$(x^2 - x + 2)^2 = x^4 + x^2 + 4 - 2x^3 + 4x^2 - 4x$$
$$= x^4 - 2x^3 + 5x^2 - 4x + 4,$$

$$(x^3 - x^2 + x - 1)^2 = x^6 + x^4 + x^2 + 1 - 2x^5 + 2x^4 - 2x^3 - 2x^3 + 2x^2 - 2x$$
$$= x^6 - 2x^5 + 3x^4 - 4x^3 + 3x^2 - 2x + 1.$$

48. *Evolution*, being an inverse process, is not so simple as involution, and the rules for it must be obtained by observing how the power of a compound quantity is formed from the quantity itself.

49. *To extract the square root of a compound quantity.*

Since the square of $a+b$ is $a^2+2ab+b^2$, we may obtain a general rule for the extraction of the square root by observing how $a+b$ may be deduced from $a^2+2ab+b^2$.

Arrange the expression according to powers (say *descending*) of some letter a. The square root of the first term a^2 is a, which is the first term in the root; subtract its square from

$$\begin{array}{r|l} & a^2 + 2ab + b^2 \ (a+b \\ & a^2 \\ \hline 2a+b) & 2ab + b^2 \\ & 2ab + b^2 \\ \hline \end{array}$$

the given expression, and bring down the remainder $2ab+b^2$; divide $2ab$ by $2a$, and the result is b, the other term in the root; multiply $2a+b$ by b and subtract the product from the remainder $2ab+b^2$. If the operation does not terminate here, *i. e.* if there is another remainder, this will shew that there are more than two terms in the root; in this case we may consider the two terms $a+b$ already found as one, and as corresponding to the term a in the preceding operation; and the square of this quantity, viz. $a^2+2ab+b^2$, having been by the preceding process subtracted from the given expression, we may divide the remainder by $2(a+b)$ for the next term in the root, and for a new subtrahend multiply $2(a+b)+$ the new term by that new term. The process may be repeated as often as necessary.

Example. Find the square root of

$$\begin{array}{rl} & x^6-2x^4+2x^3+x^2-2x+1\ (x^3-x+1 \\ & x^6 \\ \hline 2x^3-x) & -2x^4+2x^3+x^2 \\ & -2x^4+x^2 \\ \hline 2x^3-2x+1) & 2x^3-2x+1 \\ & 2x^3-2x+1 \\ \hline \end{array}$$

50. *To extract the cube root of a compound quantity.*

We have seen (Art. 46) that $(a+b)^3=a^3+3a^2b+3ab^2+b^3$. Hence we deduce the rule as in the case of the square root. Arrange the expression according to descending powers of a, the cube root of the first term a^3 is a, the first term of the root; subtract its cube from the given expression, and bring down the remainder; divide the first term by $3a^2$, the quotient is b the second

$$\begin{array}{rl} & a^3+3a^2b+3ab^2+b^3\ (a+b \\ & a^3 \\ \hline 3a^2) & 3a^2b+3ab^2+b^3 \\ & 3a^2b+3ab^2+b^3 \\ \hline \end{array}$$

term of the root; subtract the quantity $3a^2b + 3ab^2 + b^3$, if there is no remainder, the root is extracted, if there is we must proceed as before, considering $a+b$ as one term corresponding to a in the first operation.

EXAMPLE. Find the cube root of

$$x^6 + 12x^5 + 60x^4 + 160x^3 + 240x^2 + 192x + 64 \left(x^2 + 4x + 4\right.$$

$$x^6$$

$$3x^4)\ 12x^5 + 60x^4 + 160x^3$$

$$(3a^2b=)\ 12x^5$$

$$(3ab^2=)\qquad + 48x^4$$

$$(b^3=)\qquad\qquad + 64x^3$$

$$3x^4 + 24x^3 + 48x^2)\ 12x^4 + 96x^3 + 240x^2 + 192x + 64$$

$$(3a^2b=)\qquad 12x^4 + 96x^3 + 192x^2$$

$$(3ab^2=)\qquad\qquad 48x^2 + 192x$$

$$(b^3=)\qquad\qquad\qquad + 64$$

The cases of the *square* and *cube* root have been given separately on account of their more frequent occurrence, and in order to explain the rules for their extraction in arithmetic, but they are both included in the following investigation of the method of extracting the n^{th} root of a polynomial.

51. We must premise the following

LEMMA. *The first two terms of* $(a + b)^n$, *where* n *is a positive whole number, are* $a^n + na^{n-1}b$.

For by actual multiplication this is seen to be the case when $n = 2$, because $(a + b)^2 = a^2 + 2ab + b^2$, and when $n = 2$ $a^n + na^{n-1}b = a^2 + 2ab$. Now, *suppose* the proposition to be true for any value of n, i.e. *suppose* that

$$(a + b)^n = a^n + na^{n-1}b + \text{terms involving lower powers of } a,$$

$$\text{then } (a+b)^{n+1} = (a+b)\left\{a^n + na^{n-1}b + \ldots\ldots\ldots\right\}$$
$$= a^{n+1} + na^n b + \ldots\ldots\ldots$$
$$+ a^n b + \ldots\ldots\ldots$$

by actual multiplication,

$$= a^{n+1} + (n+1)a^n b + \ldots\ldots\ldots$$

which shews that *if* the Lemma be true for any value of n, it is true for the whole number next greater; but it *is* true when $n = 2$, $\therefore$ it is true when $n = 3$, $\therefore$ when $n = 4$, $\therefore$ &c. $\therefore$ generally true.

52. *To find the* nth *root of a polynomial.*

Suppose the n^{th} root to be $a + bx + cx^2 + \ldots\ldots\ldots$ which is arranged according to ascending powers of some letter x: then the given polynomial is $(a + bx + cx^2 + \ldots\ldots\ldots)^n$, which however we are to suppose expanded and arranged according to ascending powers of x.

Now by Lemma

$$(a + \overline{bx + cx^2 + \ldots\ldots})^n = a^n + na^{n-1}(bx + cx^2 + \ldots\ldots) + \ldots\ldots$$
$$= a^n + na^{n-1}bx + \text{ terms involving powers of } x \text{ above the } \textit{first}.$$

The *first* term a of the required root is known by inspection, being the n^{th} root of a^n; subtract a^n from the given expression, then the first term of the remainder is $na^{n-1}bx$; divide this by na^{n-1} and we have bx the *second* term of the root.

Again, by Lemma

$$(\overline{a + bx} + \overline{cx^2 + \ldots\ldots})^n = (a+bx)^n + n(a+bx)^{n-1}(cx^2 + \ldots\ldots) + \&c.$$
$$= (a+bx)^n + na^{n-1}cx^2 + \text{ terms involving powers of } x \text{ above the } \textit{second}.$$

The terms a and bx are already known; if then we subtract $(a + bx)^n$ from the given expression, the first term of the remainder will be $na^{n-1}cx^2$, dividing which by the

same quantity as in the first process, viz. na^{n-1}, we have cx^2 the *third* term of the root.

By precisely similar reasoning it will appear, that if we subtract $(a + bx + cx^2)^n$ from the given polynomial and divide the first term of the remainder by na^{n-1}, we shall obtain the *fourth* term of the root; and so on.

EXAMPLE. Extract the fourth root of

$$x^8 - 8x^7 + 28x^6 - 56x^5 + 70x^4 - 56x^3 + 28x^2 - 8x + 1\,(x^2 - 2x + 1$$

$$x^8$$

$$4x^6)\ -8x^7$$

$$x^8 - 8x^7 + 24x^6 - 32x^5 + 16x^4\ \{= (x^2 - 2x)^4\}$$

$$4x^6)\ 4x^6 - \&c.$$

$$x^8 - 8x^7 + 28x^6 - 56x^5 + 70x^4 - 56x^3 + 28x^2 - 8x + 1\ \{= (x^2 - 2x + 1)^4\}.$$

53. The preceding investigations are quite necessary in order to understand fully the theory of the extraction of the square and cube roots of *numbers*.

54. *On the rule of pointing in the extraction of the square root of a number.*

Every number consisting of one figure or digit is less than 10, and therefore the square of a number of one figure is less than 10^2; more generally, every number of n figures is less than 10^n, (because 10^n represents 1 followed by n cyphers,) and therefore the square of such a number is less than 10^{2n}, but also every number of n figures is not less than 10^{n-1}, and therefore its square is not less than 10^{2n-2}; now 10^{2n-2} is the smallest number of $2n-1$ figures and 10^{2n} the smallest of $2n+1$ figures, consequently the square of a number of n figures has either $2n$ or $2n-1$ figures. This being the case if we put a point over the unit's place of a number of which

the root is to be extracted, and point every second figure from right to left, the number of points will always be equal to the number of figures in the root: if the number of figures be even, the number will be divided into compartments of two each; if odd, the last compartment will contain only a single figure.

Ex. $1\dot{7}2\dot{4}3\dot{6}$, $\dot{2}1\dot{5}4\dot{7}$: each of these numbers has three figures in its square root.

The rule for extracting the square root of a number is only an adaptation of that for extracting the square root of an algebraical expression. The nature of the adaptation will be seen best by an example.

Let it be required to extract the square root of 2116.

Point the unit's place and every second figure; find the greatest number the square of which is not greater than the number expressed by the first period; in the example 21 is the first period and 4^2 is not greater than 21, hence 4 is the first figure in the root. Then subtract the square of the number thus found from the first period and bring down the second; divide this number, omitting the last figure, by twice the number already found, the quotient is the second figure of the root; in the example we divide 51 by 8 which gives 6 for the second figure. Annex the figure thus found to the divisor, and multiply the divisor so increased by the figure of the root last found to form the subtrahend; in the example 86 is multiplied by 6, which gives the subtrahend 516. If there be more periods to be brought down, the operation must be repeated.

$$
\begin{array}{rl}
 & 2\dot{1}1\dot{6}\ (46 \\
 & 16 \\
\hline
86) & 516 \\
 & 516 \\
\hline
\end{array}
$$

55. *On pointing in the extraction of the cube root.*

It may be shewn exactly in the same way as in the case of the square root, (Art. 54) that the cube of a number of n figures contains $3n$, $3n-1$, or $3n-2$ figures, and therefore that if we put a point over the unit's place and on each third figure we shall have as many periods of figures as there are figures in the root.

56. The rule for the extraction of the cube root of a number is deduced from that for the extracting of the cube root of an algebraical expression, exactly in the same way as in the case of the square root.

Let it be required to extract the cube root of 12167.

$$\begin{array}{rrl} & \dot{1}2\dot{1}67 & (\overset{a}{20} + \overset{b}{3} = 23 \\ & 8 & \\ \hline 3a^2 = 1200) & 4167 & \\ & 3600 & = 3a^2b \\ & 540 & = 3ab^2 \\ & 27 & = b^3 \\ \hline & 4167 & = \text{subtrahend.} \end{array}$$

Point the number according to the rule; in the example there are two periods. Find the greatest number the cube of which is not greater than the number expressed by the first period, this will be the first figure in the root; in the example it is 2: in order to compare this operation with the algebraical one call this figure with a cypher affixed to it a, so that in the example $a = 20$, and let b be the next figure required. Subtract the cube of the figure already found from the first period and bring down the second, divide this by $3a^2$ and the quotient will *probably* be the next figure b of the root; in the example we find $b = 3$. Form the subtrahend by compounding the formula $3a^2b + 3ab^2 + b^3$, where a and b have the values already found; if this subtrahend is too large, take the number next less than that found for b and try again, and so on until you find a subtrahend sufficiently small. If there are more than two figures in the root, bring down the next period and proceed as before; the subtrahend will always be found by the formula $3a^2b + 3ab^2 + b^3$, it being

remembered that a stands for all the figures already found with a cypher affixed; suppose for example that in any extraction the figures 274 had been found and that 6 was the next, then $a = 2740$ and $b = 6$.

57. *To explain why it is that the rule given for finding the successive figures of the cube root of a number frequently gives a number too large.*

Suppose the part of the root found to be a, and the next digit b, then the rule is to subtract a^3 and divide by $3a^2$; but the actual quantity given by this rule is

$$\frac{3a^2b + 3ab^2 + b^3}{3a^2} = b + \frac{3ab^2 + b^3}{3a^2},$$

consequently the result given by the rule differs from the required number b by the quantity $\frac{3ab^2 + b^3}{3a^2}$, which may very well be greater than 1, and if so the number given by the rule will be too great. The rule is more likely to be in error at the commencement of the operation, because then a is not so great as afterwards.

The name of *trial divisor* has been very properly assigned to $3a^2$.

58. Hence we see why in arithmetic no rule can be given for the extraction of the higher roots; for the rule for the cube root becomes, as we have seen, uncertain; and if in the case of high roots we adopted the method of Art. (52), we should find that the *trial divisor* na^{n-1} would scarcely ever give us any help in discovering the figures of the root.

59. The distinction between the algebraical and arithmetical operations will be seen at once by observing the difference in the operations of squaring (or raising to any power) an algebraical expression and a number. We have

$$(ax + b)^2 = a^2x^2 + 2abx + b^2.$$

Now suppose $x = 10$, then $10a + b$ represents a number having digits a and b, but $(10a + b)^2$ is not represented arithmetically by $10^2a^2 + 10 . 2ab + b^2$ *unless* $2ab$ *is less than* 10, and if this be the case the square of the number will have lost algebraically the type of the number itself.

For example, $18 = 10 + 8$, but 18^2 is not represented by $10^2 + 10.16 + 64$, but by $10^2.3 + 10.2 + 4$.

60. *To investigate a rule for pointing in the extraction of the square root of a decimal quantity.*

A decimal quantity may be represented by $\frac{N}{10^n}$, where N represents the number supposing the decimal point omitted, and n is the number of decimal places. Now n is either odd or even; if it is odd, multiply numerator and denominator by 10, and let the quantity thus modified be represented by $\frac{M}{10^{2m}}$.

Then $\sqrt{\frac{M}{10^{2m}}} = \frac{\sqrt{M}}{10^m}$, a formula which indicates that the square root of M is to be extracted as in whole numbers, and that m places are to be marked off for decimals; but in pointing M it is to be observed that, since we made the number of decimal places even, a point will necessarily fall on the original unit's place. Hence we have this rule: Put a point over the unit's place and point every second figure right and left.

The rule for pointing in the extraction of the cube root may be found in a similar manner.

61. *When* p *figures of a square root have been obtained by the ordinary method,* p − 1 *more may be obtained by division only.*

Let a be the part of the root already obtained, x the part consisting of $p - 1$ figures which we wish to obtain, and let N be the whole root, so that

$$N = a\,10^{p-1} + x,$$

$$\therefore\ N^2 = a^2\,10^{2p-2} + 2ax10^{p-1} + x^2.$$

3—2

Subtracting $a^2 10^{2p-2}$ from each side of this equation and dividing by $2a10^{p-1}$ we have

$$\frac{N^2 - a^2 10^{2p-2}}{2a10^{p-1}} = x + \frac{x^2}{2a10^{p-1}}.$$

From this it appears that the division above indicated will give us x correctly, if we can prove $\frac{x^2}{2a10^{p-1}}$ to be a proper fraction.

Now x consists of $p-1$ figures, and $\therefore$ is $< 10^{p-1}$;

$$\therefore\ x^2 \text{ is } < 10^{2p-2};$$

but a consists of p figures, and $\therefore$ is $> 10^{p-1}$;

$$\therefore\ \frac{x^2}{a10^{p-1}} < \frac{10^{2p-2}}{10^{2p-2}} < 1;$$

$$\therefore\ \frac{x^2}{2a10^{p-1}} < \frac{1}{2},$$

and hence the division will give us the $p-1$ figures of x correctly.

ON EQUATIONS.

62. An *equation* has already been defined to be an algebraical sentence expressing the equality of two algebraical expressions, or (which is the same thing) of an algebraical expression to zero.

If an unknown quantity is involved, the equation involving it serves to determine the unknown quantity, and it is our business now to lay down rules for the performance of this process, which is called the *solution* of the equation.

If when cleared of radicals an equation involves only the *first* power of the unknown quantity x, it is called a *simple* equation; if it involves x^2 also, it is called a *quadratic;* and, generally, if it involves x^n it is said to be *an equation of n*

dimensions. We shall in this treatise be concerned only with *simple* and *quadratic* equations.

A value of x which satisfies an equation is called a *root* of the equation.

63. Suppose we have the equation $x + a = b$; then subtracting a from each side we have $x = b - a$, that is, a quantity may be placed on the other side of an equation if its sign be changed. This process, which is one of the most frequent in the solution of equations, is called *transposition*.

64. It is manifest that if the same operation be performed on the two sides of an equation, the equality will still subsist; we may therefore multiply or divide the two sides of an equation by the same quantity, or may raise the two sides to the same power or extract any root of both sides. If for example

$$x^2 + 2x - 3 = 3x^2 + 1,$$

then will the following equations hold good,

$$P(x^2 + 2x - 3) = P(3x^2 + 1),$$

$$\frac{x^2 + 2x - 3}{P} = \frac{3x^2 + 1}{P},$$

$$(x^2 + 2x - 3)^n = (3x^2 + 1)^n$$

$$\sqrt[n]{x^2 + 2x - 3} = \sqrt[n]{3x^2 + 1}.$$

65. A frequent process in the solution of equations is *clearing the equation of radicals;* this is done by putting any radical of which we desire to rid the equation on one side by itself, and transposing all the other terms to the other side, we then raise both sides of the equation to the power indicated by the radical, which consequently disappears. The process will be seen best by an example: suppose

$$\sqrt{x + 1} + \sqrt{x - 1} = 2.$$

The process of clearing this equation of radicals will stand as follows,

transposing, $\sqrt{x+1} = 2 - \sqrt{x-1}$,

squaring, $x + 1 = 4 - 4\sqrt{x-1} + x - 1$,

transposing, $4\sqrt{x-1} = 2$,

dividing by 4, $\sqrt{x-1} = \frac{1}{2}$,

squaring, $x - 1 = \frac{1}{4}$;

which is a simple equation free from radicals.

It is obvious that it is generally impossible to ascertain the *degree* of an equation, that is, whether it is simple, quadratic or of any higher degree, until it has been cleared of radicals.

66. An equation is cleared of fractions by multiplying both sides of it by the least common multiple of the denominators. It is however perhaps practically the easiest method to multiply by each of the denominators in succession, and make such simplifications as the case allows after each multiplication.

ON SIMPLE EQUATIONS.

67. *To find the value of an unknown quantity in a simple equation.*

An equation given as a simple equation may involve radicals; if so, let them be got rid of first. Next, clear the equation of fractions. Next, transpose all terms involving the unknown quantity to the left-hand side, and all terms involving only known quantities to the right-hand side of the equation. Divide both sides by the coefficient, or sum of the coefficients, of the unknown quantity, and the value required is obtained.

Ex. 1.
$$2x + 3 = 3x - 1,$$
$$3x - 2x = 3 + 1,$$
$$x = 4.$$

Ex. 2.
$$\frac{2x - 1}{3} + \frac{3x - 2}{2} = \frac{5}{6},$$
$$2(2x - 1) + 3(3x - 2) = 5,$$
$$4x + 9x = 2 + 6 + 5,$$
$$13x = 13,$$
$$x = 1.$$

Ex. 3.
$$\sqrt{x + 1} + \sqrt{x - 1} = 2,$$
$$\sqrt{x + 1} - 2 = -\sqrt{x - 1},$$
$$x + 1 + 4 - 4\sqrt{x + 1} = x - 1,$$
$$4\sqrt{x + 1} = 6,$$
$$\sqrt{x + 1} = \frac{3}{2},$$
$$x + 1 = \frac{9}{4},$$
$$x = \frac{9}{4} - 1 = \frac{5}{4}.$$

ON QUADRATIC EQUATIONS.

68. All the rules for the simplification of equations which have already been given apply equally to quadratic equations; indeed, as has been observed, it is not always, until the simplification has been effected, that we are able to say whether the equation is simple or quadratic. By such sim-

plification the equation, if a quadratic, will be reduced to one of these forms,

$$x^2 = b,$$

$$\text{or } x^2 + ax = b.$$

In the *first* case we have at once by extracting the square root of both sides

$$x = \pm\sqrt{b}.$$

Let the student take particular notice of the double sign prefixed to the radical, and *which in the solution of quadratic equations ought always to be prefixed*, because $-\sqrt{b}$ and $+\sqrt{b}$ satisfy the equation $x^2 = b$ equally well, since the square of either of them is $+b$.

In the *second* case the solution is not so obvious, but it is easily effected, by observing that the quantity $\frac{a^2}{4}$ added to each side of the equation will make the left-hand side a complete square; in fact, we have

$$x^2 + ax + \frac{a^2}{4} = b + \frac{a^2}{4},$$

which may be written thus

$$\left(x + \frac{a}{2}\right)^2 = b + \frac{a^2}{4},$$

extracting the square root of both sides we have

$$x + \frac{a}{2} = \pm\sqrt{b + \frac{a^2}{4}},$$

transposing, $$x = -\frac{a}{2} \pm \sqrt{b + \frac{a^2}{4}}.$$

The two values of x,

$$-\frac{a}{2} + \sqrt{b + \frac{a^2}{4}} \text{ and } -\frac{a}{2} - \sqrt{b + \frac{a^2}{4}},$$

satisfy the equation $x^2 + ax = b$, and are called its roots.

The student may, if he pleases, actually substitute either of the expressions which we have found in the equation, and he will find the result to be an identity.

69. Hence, further, we see that every quadratic has *two roots* and no more, and in solving quadratics the student should be careful always to represent both roots.

70. The following form of a quadratic is included in the preceding, by supposing b negative, but it is worthy of separate consideration. Suppose

$$x^2 + ax = -b.$$

Completing the square as before,

$$x^2 + ax + \frac{a^2}{4} = \frac{a^2}{4} - b,$$

$$x + \frac{a}{2} = \pm\sqrt{\frac{a^2}{4} - b},$$

$$x = -\frac{a}{2} \pm \sqrt{\frac{a^2}{4} - b}.$$

Now, suppose that $\frac{a^2}{4}$ is less than b, then $\frac{a^2}{4} - b$ is a negative quantity, and the expression $\sqrt{\frac{a^2}{4} - b}$ represents an operation *which cannot be performed*, for there is no quantity the square of which is negative. Quantities of this kind are called *impossible* or *imaginary*, and the roots of an equation when they involve such quantities are called *impossible* or *imaginary* roots; so far as we are concerned at present however roots of this kind are quite as much the object of our search as *real* roots, since the question is merely this, what symbolical quantity substituted for x and operated upon according to the rules of algebra will satisfy a given equation?

It is manifest that if one root of a quadratic is imaginary the other is also imaginary.

71. An equation may sometimes be treated as a quadratic which has higher powers of x in it than the second; in fact every equation is virtually a quadratic which has only two powers of x involved, one of which is twice as great as the other. See Ex. (4).

Ex. 1.
$$2x^2 + 3x + 1 = x^2 + 5,$$
$$x^2 + 3x = 4,$$
$$x^2 + 3x + \frac{9}{4} = \frac{9}{4} + 4 = \frac{25}{4},$$
$$x + \frac{3}{2} = \pm \frac{5}{2},$$
$$x = \frac{-3 \pm 5}{2} = 1 \text{ or } -4.$$

Ex. 2.
$$\frac{2}{x-1} + \frac{1}{x} = 4,$$
$$2x + x - 1 = 4x(x-1),$$
$$4x^2 - 4x - 3x = -1,$$
$$4x^2 - 7x = -1,$$
$$x^2 - \frac{7x}{4} = -\frac{1}{4},$$
$$x^2 - \frac{7x}{4} + \frac{49}{64} = \frac{49}{64} - \frac{16}{64} = \frac{33}{64},$$
$$x = \frac{7 \pm \sqrt{33}}{8}.$$

Ex. 3.
$$x^2 + 2x + 3 = 4x + 1,$$
$$x^2 - 2x = -2,$$
$$x^2 - 2x + 1 = -1,$$
$$x - 1 = \pm \sqrt{-1},$$
$$x = 1 \pm \sqrt{-1}.$$

Ex. 4.

$$x^6 + 2x^3 = 8,$$
$$x^6 + 2x^3 + 1 = 9,$$
$$x^3 + 1 = \pm 3,$$
$$x^3 = -1 \pm 3 = 2 \text{ or } -4,$$
$$x = \sqrt[3]{2} \text{ or } -\sqrt[3]{4}.$$

ON SIMULTANEOUS EQUATIONS.

72. We have seen how it is possible to find the value of x which satisfies a given simple or quadratic equation; but sometimes the problem is presented of finding *two unknown quantities from two equations;* two equations which are thus given to determine two quantities x and y, involved in both, are said to be *simultaneous*.

73. The simple rule for the solution of such equations is to find the value of one of the unknown quantities (y), in terms of the other (x) from one equation, and substitute the value so found, in the other; we shall thus have an equation involving x only for determining x, and this may be a simple equation or a quadratic, according to circumstances.

74. The process just described is not always in practice the most convenient; it is manifest that it does not signify in what manner the quantity y is got rid of between the two equations, and we may therefore give this rule: *Eliminate* (*i. e.* get rid of) y between the two equations, and obtain x from the result. The ingenuity of the student will frequently be exercised in determining the most convenient mode of elimination.

75. It is not difficult to see that *two* equations, and no more, are necessary for the determination of *two* unknown quantities; in like manner, three equations will be necessary and sufficient to determine three unknown quantities, and so on. The name *simultaneous* is applied to any such system of equations, however many there may be.

Ex. 1. Given
$$\left.\begin{aligned} ax + by &= c \quad (1) \\ a'x + b'y &= c' \quad (2) \end{aligned}\right\}$$
to find x and y.

Multiply (1) by b', and (2) by b, and the equations become

$$ab'x + bb'y = b'c,$$

$$a'bx + bb'y = bc'.$$

Subtract one of these from the other, and there results

$$(ab' - a'b)\,x = b'c - bc';$$

$$\therefore\ x = \frac{b'c - bc'}{ab' - a'b}.$$

To find y we have from (1)

$$y = \frac{1}{b}(c - ax),$$

writing for x its value

$$y = \frac{1}{b}\left(c - a\frac{b'c - bc'}{ab' - a'b}\right)$$

$$= \frac{1}{b}\cdot\frac{abc' - a'bc}{ab' - a'b} = \frac{ac' - a'c}{ab' - a'b}.$$

Thus we have found both x and y; the latter however might have been determined more neatly by treating (1) and (2) as we did in finding x, that is, multiplying the former by a', the latter by a, and subtracting the resulting equations.

76. It may perhaps be also worth while to remark even in this early example that, in a system of equations such as (1) and (2), when x has been found, y may be known by inspection. For it is to be observed, that in (1) and (2) x bears the same relation to a and a', that y bears to b and b'; in fact, if in those equations we write b for a, and b' for a', and lastly, interchange x and y, the equations remain unchanged; hence we conclude, that the value of y may be obtained from that of x by writing b for a, b' for a', and of course a for b, and a' for b'.

Now $$x = \frac{b'c - bc'}{ab' - a'b};$$

$$\therefore\ y = \frac{a'c - ac'}{ba' - b'a},$$

which is the same result as before, though written in a manner slightly different.

Ex. 2.
$$\left.\begin{aligned} \frac{x-1}{2} + \frac{2y-1}{3} &= 1 \qquad (1) \\ 2x + 1 - \frac{y+2}{3} &= 0 \qquad (2) \end{aligned}\right\}.$$

These equations must first be cleared of fractions and put in their simplest form.

Multiplying (1) by 6, there results

$$3x - 3 + 4y - 2 = 6,$$

$$\text{or } 3x + 4y = 11 \qquad (3).$$

Multiplying (2) by 3,

$$6x + 3 - y - 2 = 0,$$

$$\text{or } 6x - y + 1 = 0;$$

$$\therefore\ y = 6x + 1 \qquad (4).$$

Writing this value for y in (3),

$$3x + 4(6x + 1) = 11,$$

$$27x = 11 - 4 = 7;$$

$$\therefore\ x = \frac{7}{27};$$

and therefore from (4)

$$y = 6 \times \frac{7}{27} + 1 = \frac{14}{9} + 1 = \frac{23}{9}.$$

Ex. 3.
$$\left.\begin{aligned} x + y &= a \qquad (1) \\ xy &= b \qquad (2) \end{aligned}\right\}.$$

From (1) $\qquad y = a - x,$

putting this value for y in (2), we have

$$x(a - x) = b,$$

$$\text{or } x^2 - ax = -b.$$

Completing the square according to the rule of Art. (68),

$$x^2 - ax + \frac{a^2}{4} = \frac{a^2}{4} - b,$$

$$x - \frac{a}{2} = \pm \sqrt{\frac{a^2}{4} - b},$$

$$x = \frac{a}{2} \pm \sqrt{\frac{a^2}{4} - b},$$

and $y = a - x = \frac{a}{2} \mp \sqrt{\frac{a^2}{4} - b}$.

Ex. 4.
$$x + y + z = 0 \qquad (1),$$
$$x + 2y + 3z = 1 \qquad (2),$$
$$2x + y + 3z = 2 \qquad (3).$$

Subtracting (1) from (2),

$$y + 2z = 1 \qquad (4).$$

Multiplying (1) by 2, and subtracting the result from (3),

$$-y + z = 2 \qquad (5).$$

Adding (4) and (5),

$$3z = 3,$$
$$z = 1;$$

therefore from (4) $\quad y = 1 - 2z = 1 - 2 = -1,$

and from (1) $\quad x = -y - z = 1 - 1 = 0.$

Ex. 5.
$$x + y + z = 1 \qquad (1),$$
$$(b + c)\,x + (a + c)\,y + (a + b)\,z = 0 \qquad (2),$$
$$bcx + acy + abz = 0 \qquad (3).$$

Multiply (2) by bc and (3) by $b + c$ and subtract, then

$$\{(a + c)\,bc - (b + c)\,ac\}\,y + \{(a + b)\,bc - (b + c)\,ab\}\,z = 0,$$

or $(bc^2 - ac^2)\,y + (b^2c - ab^2)\,z = 0$;

$$\therefore\ y = -\frac{b^2}{c^2}\,\frac{c - a}{b - a}\,z.$$

Similarly we should find (see the remarks on Example 1),

$$x = -\frac{a^2}{c^2}\frac{c-b}{a-b}z;$$

therefore substituting in (1)

$$-\frac{a^2}{c^2}\frac{c-b}{a-b}z - \frac{b^2}{c^2}\frac{c-a}{b-a}z + z = 1,$$

$$\text{or } z\left(1 + \frac{a^2}{c^2}\frac{b-c}{a-b} + \frac{b^2}{c^2}\frac{c-a}{a-b}\right) = 1,$$

$$z\left\{c^2(a-b) + a^2(b-c) + b^2(c-a)\right\} = c^2(a-b),$$

$$z\left\{c^2(a-b) + ab(a-b) - c(a^2-b^2)\right\} = c^2(a-b).$$

Dividing by $a-b$

$$z\left\{c^2 + ab - c(a+b)\right\} = c^2,$$

$$\text{or } z(c-a)(c-b) = c^2;$$

$$\therefore z = \frac{c^2}{(c-a)(c-b)}.$$

In like manner $y = \dfrac{b^2}{(b-a)(b-c)}$,

and $x = \dfrac{a^2}{(a-b)(a-c)}$.

77. Sometimes a system of equations may be given which are not really sufficient to determine the unknown quantities, in consequence of not being *independent*, that is to say, of any one of the equations being deducible from the rest. This dependence is not always to be detected by inspection, but will become apparent if we endeavour to solve the equations.

Ex. Let it be required to determine x, y and z from the following system:

$$2x + 3y + z = 11 \quad (1),$$

$$x - y + 2z = 5 \quad (2),$$

$$x + 9y - 4z = 7 \quad (3).$$

Multiplying (2) by 3, and adding the result to (1), we have

$$5x + 7z = 26 \qquad (4),$$

Again, multiplying (2) by 9 and adding (3),

$$10x + 14z = 52 \qquad (5).$$

The two equations which we have thus obtained, viz. (4) and (5) are identical, and therefore the given system is not sufficient to determine x, y and z; in fact those three equations are equivalent to only *two* independent equations, any one of them being derivable from the other two.

78. The preceding examples must serve for the elucidation of the method of solving equations; the illustrations of the process might be indefinitely extended, but a familiar acquaintance with the most convenient methods can only be acquired by the practice of actual solution on the part of the student himself.

ON PROBLEMS WHICH MAY BE RESOLVED BY MEANS OF ALGEBRAIC EQUATIONS.

79. A vast variety of questions, which present great and perhaps insuperable difficulty to a mind unaided by the art of symbolical reasoning, are rendered extremely simple by reducing them to algebraic equations.

80. The most general rule which can be given for the solution of such questions is this: Denote the unknown quantities of the problem by symbols (x, y, z, &c.) and then express the conditions of the problem in terms of those symbols; we do by this means, in fact, express by algebraical sentences or equations the ordinary written sentences in which the problem is given. The equations thus constructed must be solved according to the methods which have been previously discussed; the equations may, for aught we can tell, by inspection of the problem *a priori*, rise to a degree above the second, but of course we shall confine our attention here to those problems which produce either simple equations or quadratics.

Ex. 1. Divide the number 16 into two parts such that their difference shall be equal to half the number itself.

Let x represent one of the parts, then will $16 - x$ represent the other and $16 - 2x$ will represent the difference of the two; but by the question this difference is equal to $\frac{16}{2}$ or 8;

$$\therefore\ 16 - 2x = 8,$$

$$2x = 8,$$

$$x = 4 \quad \text{one of the parts,}$$

$$16 - x = 12 \quad \text{the other part.}$$

Ex. 2. Two pipes will separately fill a cistern in a hours and b hours respectively; in how long a time will they fill it together?

Let x be the number of hours required.

Call the whole content of the cistern 1, then since the first pipe pours in 1 (or the whole quantity required to fill the cistern) in a hours, it pours in $\frac{1}{a}$ in one hour, and therefore $\frac{x}{a}$ in x hours.

Similarly, the second pipe pours in $\frac{x}{b}$ in the same time;

therefore they together pour in $\frac{x}{a} + \frac{x}{b}$.

But this quantity is the whole content of the cistern or 1;

$$\therefore\ \frac{x}{a} + \frac{x}{b} = 1,$$

$$\text{and } x = \frac{1}{\frac{1}{a} + \frac{1}{b}} = \frac{ab}{a+b}$$

Ex. 3. A's money exceeds B's and C's by £a and £b respectively, and that of B and C together is £c: find the sum possessed by each.

Let $x =$ the sum possessed by A;

$\therefore\ x - a =$ .... B,

and $x - b =$ .... C;

therefore by question,

$$x - a + x - b = c,$$

$$2x = a + b + c,$$

$$x = \frac{a+b+c}{2} = A\text{'s},$$

$$x - a = \frac{-a+b+c}{2} = B\text{'s},$$

$$x - b = \frac{a-b+c}{2} = C\text{'s}.$$

Ex. 4. Divide £a among three persons, so that the first may have m times as much as the second, and the third n times as much as the first and second together.

Let $x =$ the sum allotted to the second;

$\therefore\ mx =$ .... first,

and $n(x + mx) =$ .... third;

$$\therefore\ x + mx + n(x + mx) = a,$$

$$x(1 + m)(1 + n) = a,$$

$$x = \frac{a}{(1+m)(1+n)},$$

$$mx = \frac{ma}{(1+m).(1+n)},$$

$$n(x + mx) = \frac{na}{1+n}.$$

Ex. 5. Divide the quantity a into two parts such that the product of the whole and one of the parts shall be equal to the square of the other part.

$$\text{Let } x = \text{one part;}$$

$$\therefore\ a - x = \text{the other;}$$

$\therefore$ by the question,

$$ax = (a - x)^2$$

$$= a^2 - 2ax + x^2,$$

$$x^2 - 3ax = -a^2;$$

completing the square,

$$x^2 - 3ax + \frac{9a^2}{4} = \frac{9a^2}{4} - a^2 = \frac{5a^2}{4};$$

$$\therefore\ x - \frac{3a}{2} = \pm \frac{\sqrt{5}}{2}\,a,$$

$$x = \frac{3 \pm \sqrt{5}}{2}\,a.$$

We must take the negative sign, because $\frac{3 + \sqrt{5}}{2}a$ is greater than a;

$$\therefore\ x = \frac{3 - \sqrt{5}}{2}\,a,$$

$$a - x = \frac{\sqrt{5} - 1}{2}\,a.$$

Ex. 6. The sum of two numbers is a, and the sum of their cubes b; find the numbers.

$$\text{Let } x = \text{one of the numbers;}$$

$$\therefore\ a - x = \text{the other;}$$

$\therefore$ by question,

$$x^3 + (a - x)^3 = b,$$

$$\text{or } x^3 + a^3 - 3a^2x + 3ax^2 - x^3 = b,$$

$$x^2 - ax = \frac{b - a^3}{3a},$$

$$x^2 - ax + \frac{a^2}{4} = \frac{b - a^3}{3a} + \frac{a^2}{4};$$

$$\therefore\ x = \frac{a}{2} \pm \sqrt{\frac{b - a^3}{3a} + \frac{a^2}{4}},$$

$$a - x = \frac{a}{2} \mp \sqrt{\frac{b - a^3}{3a} + \frac{a^2}{4}}.$$

Hence the two numbers are $\frac{a}{2} + \sqrt{\frac{b - a^3}{3a} + \frac{a^2}{4}}$,

and $\frac{a}{2} - \sqrt{\frac{b - a^3}{3a} + \frac{a^2}{4}}$.

Ex. 7. There are three magnitudes, the sum of the first and second of which is a, that of the first and third b, and that of the second and third c; find them.

Let the magnitudes be represented by x, y, z respectively; then,

$$x + y = a,$$

$$x + z = b,$$

$$y + z = c;$$

adding these equations and dividing the result by 2,

$$x + y + z = \frac{a + b + c}{2}.$$

Hence $x = \frac{a + b + c}{2} - c = \frac{a + b - c}{2}$,

$$y = \frac{a + b + c}{2} - b = \frac{a - b + c}{2},$$

$$z = \frac{a + b + c}{2} - a = \frac{-a + b + c}{2}.$$

Ex. 8. Required four magnitudes the products of which taken three together are a^3, b^3, c^3 and d^3.

Call the magnitudes x, y, z and u.

Then
$$yzu = a^3,$$
$$xzu = b^3,$$
$$xyu = c^3,$$
$$xyz = d^3.$$

Multiplying these equations together, we have
$$x^3y^3z^3u^3 = a^3b^3c^3d^3;$$
$$\therefore\ xyzu = abcd;$$
$$\therefore\ x = \frac{abcd}{yzu} = \frac{abcd}{a^3},$$
$$y = \frac{abcd}{b^3},$$
$$z = \frac{abcd}{c^3},$$
$$u = \frac{abcd}{d^3}.$$

ON RATIOS.

81. Ratio is the relation which quantities of the same kind bear to each other in respect of magnitude.

Thus 6 is twice as great as 3, and 2 is twice as great as 1; therefore we should say that the ratio of 6 to 3 is the same as that of 2 to 1, or as we may write for shortness' sake,

$$6 : 3 :: 2 : 1.$$

In speaking of the ratio of two quantities $a : b$, a and b are called the *terms* of the ratio, and a is distinguished as the *antecedent*, b as the *consequent*.

82. It is easy to shew that the terms of a ratio may be multiplied or divided by any (the same) number. For distinctness' sake let us consider a and b as representing two *lines*, then by the symbol a we mean to denote a line a times as great as a certain standard line (a *foot* for instance), and by b a line b times as great. The lines in question are then in the proportion of the numbers a, b; but if we had taken a line only half as long (six inches) for the standard, the lines would have been represented by $2a$, and $2b$; but their ratio of course is not altered;

$$\therefore\ a : b :: 2a : 2b.$$

In like manner it would appear that $a : b :: 3a : 3b$, and generally that the terms of a ratio may be multiplied by any number without affecting the value of the ratio.

Conversely, the terms may be *divided* by any number.

83. *Hence it follows that we may represent ratios algebraically by fractions, of which the antecedent is the numerator and the consequent the denominator.*

For by what has been said the ratio of $a : b$ is the same as the ratio of $\frac{a}{b} : 1$; now 1 is a given quantity, and therefore we may take the fraction $\frac{a}{b}$ as the symbol of the ratio $\frac{a}{b} : 1$, and therefore as the representative of the ratio $a : b$.

In fact, the ratio $a : b$ may be conceived to mean that a is as many times as great as b, as $\frac{a}{b}$ is as great as 1, and therefore the magnitude of the fraction $\frac{a}{b}$ *measures* the magnitude of the ratio of $a : b$.

Henceforth therefore we shall represent the ratio of $a : b$ by the fraction $\frac{a}{b}$.

If a is greater than b, the ratio $\frac{a}{b}$ is called a ratio of *greater inequality.*

If $a = b$, the ratio is called a ratio of *equality.*

If a is less than b, a ratio of *less inequality.*

84. *A ratio of greater inequality is diminished, and of less inequality increased, by adding the same quantity to each of its terms.*

Let $\frac{a}{b}$ be any ratio, and let x be added to each of its terms; then

$$\frac{a+x}{b+x} \text{ is } > \text{ or } < \frac{a}{b},$$

according as

$$(a+x)b \text{ is } > \text{ or } < a(b+x),$$
$$\text{or as } bx \text{ is } > \text{ or } < ax,$$
$$\text{or as } b \text{ is } > \text{ or } < a,$$

i. e. as the ratio is one of less or greater inequality.

85. Ratios are compounded by multiplying together their corresponding terms.

Thus $\frac{a}{b}$ compounded with $\frac{c}{d}$ becomes $\frac{ac}{bd}$.

86. *A ratio is increased by being compounded with another of greater inequality, and diminished by being compounded with one of less.*

For $\frac{a}{b}$ compounded with $\frac{c}{d}$ becomes $\frac{ac}{bd}$;

$$\text{but } \frac{ac}{bd} \text{ is } > \text{ or } < \frac{a}{b},$$

$$\text{according as } \frac{c}{d} \text{ is } > \text{ or } < 1,$$

$$\text{or } c \text{ is } > \text{ or } < d,$$

which proves the proposition.

ON PROPORTION.

87. Proportion is the equality of ratios; and therefore algebraically four quantities are said to be proportional, when the fraction expressing the ratio of the first and second is equal to that expressing the ratio of the third and fourth; that is, $a : b :: c : d$ when $\frac{a}{b} = \frac{c}{d}$.

88. *If* a : b :: c : d, *then* ad = bc.

For, as we have just seen,

$$\frac{a}{b} = \frac{c}{d},$$

and $\therefore\ ad = bc$.

89 *If* a : b :: c : d, *then* a ± b : b :: c ± d : d.

$$\text{For, } \frac{a}{b} = \frac{c}{d};$$

$$\therefore\ \frac{a}{b} \pm 1 = \frac{c}{d} \pm 1,$$

$$\text{or } \frac{a \pm b}{b} = \frac{c \pm d}{d}.$$

90. *If* a : b :: c : d, *then*

ma ± nb : pc ± qb :: mc ± nd : pc ± qd.

$$\text{For, } \frac{a}{b} = \frac{c}{d};$$

$$\therefore\ \frac{ma}{nb} = \frac{mc}{nd},$$

$$\frac{ma}{nb} \pm 1 = \frac{mc}{nd} \pm 1,$$

$$\frac{ma \pm nb}{nb} = \frac{mc \pm nd}{nd},$$

$$\frac{ma \pm ab}{mc \pm nd} = \frac{nb}{nd} = \frac{b}{d};$$

in like manner it may be shewn that

$$\frac{pa \pm qb}{pc \pm qd} = \frac{b}{d};$$

$$\therefore \frac{ma \pm nb}{mc \pm nd} = \frac{pa \pm qb}{pc \pm qd}.$$

91. *If* a : b :: c : d :: e : f :: &c., then

a : b :: a + c + e + ... : b + d + f + ...

For $\frac{a}{b} = \frac{c}{d} = \frac{e}{f} = \ldots;$

$$\therefore ad = bc,$$

$$af = be,$$

$$\&c. = \&c.$$

$\therefore$ by addition,

$$ad + af + \ldots = bc + be + \ldots,$$

$$ab + ad + af + \ldots = ba + bc + be + \ldots,$$

$$a\{b + d + f + \ldots\} = b\{a + c + e + \ldots\};$$

$$\therefore \frac{a}{b} = \frac{a + c + e + \ldots\ldots}{b + d + f + \ldots\ldots}.$$

92. A variety of other propositions in proportion may be demonstrated in like manner as the preceding. The greatest simplicity is introduced by this method of representing ratios by fractions, and it will be instructive to inquire into the reason of the much more complicated processes which Euclid has found it necessary to employ in the fifth book of his elements.

Euclid's definition of proportion is this; that four quantities are said to be proportional when any equimultiples whatever being taken of the first and third, and any whatever of the second and fourth; if the multiple of the first is greater than that of the second, the multiple of the third is greater than that of the fourth, if equal equal, and if less less. Now this definition is an immediate consequence of the algebraical representation of ratio; for suppose $a : b :: c : d$,

$$\text{then } \frac{a}{b} = \frac{c}{d},$$

$$\text{and } \frac{pa}{qb} = \frac{pc}{qd},$$

$$\text{and if } pa \text{ is } > qb, \ pc \text{ is } > qd,$$

$$\text{if } pa = qb, \ pc = qd,$$

$$\text{and if } pa < qb, \ pc < qd,$$

which is Euclid's definition.

And conversely, from Euclid's definition may be deduced the algebraical rule of proportion; that is, we can shew that if, by that definition, $a : b :: c : d$, then must $\frac{a}{b} = \frac{c}{d}$.

For let a, b, c, d be four quantities such that if equimultiples pa, pc be taken of a and c, and equimultiples qb, qd of b and d, if pa be $> qb$, pc is $> qd$, if equal equal, and if less less.

Then since we may choose p and q as we please, we can make $\frac{pa}{qb}$ as nearly equal to 1 as we please; we cannot, it is true, always make it precisely equal to 1, because p and q are to be whole numbers, and a and b need not be so; but nevertheless we can make the fraction $\frac{pa}{qb}$ *as near to unity as we please;* we may therefore suppose that we have taken p and q such that

$$\frac{pa}{qb} = 1;$$

but by definition we must in this case have also

$$\frac{pc}{qd} = 1;$$

$$\therefore \frac{pa}{qb} = \frac{pc}{qd},$$

$$\text{or } \frac{a}{b} = \frac{c}{d}.$$

Hence it appears that Euclid's definition of proportion and the algebraical follow each from the other; but wherein is the propriety of Euclid's peculiar definition? in this, that the algebraical test is not applicable to geometrical quantities; we can represent addition and subtraction geometrically, but not *division*, and therefore it is necessary in geometrical investigations to adopt some definition which involves only the notion of addition and of a comparison of magnitudes with reference to *greater* or *less*.

ON VARIATION.

93. When one quantity y depends upon another x, in such a manner that, if x is changed in value, the value of y is changed in the same proportion, then y is said to *vary directly* as x, or shortly, to *vary* as x.

For instance, we know by Euclid, VI. 1, that if we double the base of a triangle the vertex remaining the same we double the area, and that in whatever proportion we alter the base the area is altered in the same proportion, hence we should say that (the altitude being given) the area of a triangle *varies as the base*.

The phrase y varies as x is written thus, $y \propto x$.

94. It will be seen that we have here introduced the notion of quantities entirely different from those hitherto con-

sidered; hitherto we have had to do only with quantities which have some determinate value, but the relation between y and x implied in the fact of y varying as x does not determine either x or y, but only a relation between them. Quantities of this kind we call *variable* quantities, to distinguish them from others the value of which is determinate and which we call *constant*.

95. The relation expressed by $y \propto x$ is equivalent to the equation $y = Cx$, where C is some constant quantity; for $\frac{y}{x}$ is the ratio of y to x, and the preceding equation expresses that this is constant or always the same whatever values x and y may have, and this is the same thing as saying that when one is increased the other is increased in the same proportion.

96. If we have any two corresponding values of x and y given we can determine the quantity C; thus, suppose $y \propto x$ and it is given that when $x = 1$, $y = 2$, then we have

$$y = Cx,$$

$$\text{but } 2 = C1,$$

$$\therefore\ y = 2x.$$

97. When two quantities are connected by the relation $y = \frac{C}{x}$, y is said to vary *inversely* as x.

And when three quantities x, y, z are connected in such a manner that $z = Cxy$, z is said to vary jointly as x and y.

98. *If* $y \propto x$ *and* $z \propto y$, *then* $z \propto x$.

For let
$$y = Cx,$$

$$z = C'y;$$

$$\therefore\ z = CC'.x,$$

and CC' is constant; $\therefore\ z \propto x$.

99. *If* y ∝ x *and* z *also* ∝ x, *then* $\sqrt{yz} \propto x$.

For let
$$y = Cx,$$
$$z = C'x;$$
$$\therefore\ yz = CC' \cdot x^2,$$
$$\sqrt{yz} = \sqrt{CC'} \cdot x,$$
and $\sqrt{CC'}$ is constant; $\therefore\ \sqrt{yz} \propto x$.

Many other propositions may be demonstrated in like manner with perfect facility. We shall conclude with the following proposition.

100. *If* z *be a quantity depending upon two others,* x *and* y, *in such a manner that when* x *is constant and* y *allowed to vary* z ∝ y, *and when* y *is constant and* x *allowed to vary* z ∝ x, *then when* x *and* y *both vary* z *will* ∝ xy.

Let $z = u \cdot xy$, where u is a quantity which, for anything we know at present to the contrary, may involve x or y or both.

Then when x is constant and y variable $z \propto y$, but $z = ux \cdot y$, therefore ux does not involve y, or u does not involve y.

In like manner u does not involve x, therefore it is constant, or $z \propto xy$.

We may illustrate the preceding proposition as follows. When the base of a triangle is given the area ∝ the altitude, and when the altitude is given the area ∝ the base; hence when neither is given, the area ∝ base × altitude.

ON ARITHMETICAL PROGRESSION.

101. Def. Quantities are said to be in arithmetical progression, when they increase or decrease by a common difference.

Thus $a,\ a+d,\ a+2d, \ldots\ldots\ldots$ is an arithmetical series.

102. *To sum an arithmetical series.*

Let a be the first term, d the common difference of the terms; then the second term will be $a+d$, the third $a+2d$, and generally the n^{th} term will be $a+(n-1).d$.

Let S be the sum of n terms, then

$$S = a + a + d + a + 2d + \ldots\ldots\ldots + a + (n-1).d;$$

writing the terms in the reverse order, we have

$$S = a + (n-1)d + a + (n-2)d + a + (n-3)d + \ldots\ldots + a;$$

adding together these two equations,

$$2S = 2a + (n-1)d + 2a + (n-1)d + 2a + (n-1)d \ldots\ldots\ldots$$
$$+ 2a + (n-1)d,$$

$$= \{2a + (n-1)d\}n, \text{ since there are } n \text{ terms;}$$

$$\therefore S = \{2a + (n-1)d\}\frac{n}{2},$$

which is the expression for the sum required.

The expression for S may also be written thus: let l be the *last* term, i.e. $l = a + (n-1)d$, then

$$S = (a+l)\frac{n}{2}.$$

Cor. Any three of the quantities a, d, n and S, being given, the fourth may be found.

Ex. 1. Find the sum of 10 terms of the arithmetical series 2, 5, 8, ...

Here $a = 2, \; d = 3, \; n = 10;$

$$\therefore \; S = (4 + 9 \times 3)\, 5 = 31 \times 5 = 155.$$

Ex. 2. There is an arithmetical series the fourth term of which is 9 and the seventh 15: find the series.

The formula for the n^{th} term is $a + (n - 1)\, d$, therefore in this example we have

$$a + 3d = 9,$$

$$a + 6d = 15.$$

In subtracting the first of these equations from the second,

$$3d = 6,$$

$$d = 2;$$

$$\therefore \; a = 9 - 6 = 3;$$

$\therefore$ the series is 3, 5, 7, 9..........

Ex. 3. Insert n arithmetical means between a and b. This is in other words to form an arithmetical series of $n + 2$ terms, of which the first shall be a and the last b.

Let d be the common difference, then we must have

$$a + (n + 1)\, d = b;$$

$$\therefore \; d = \frac{b - a}{n + 1}.$$

Hence the terms required will be

$$a + \frac{b - a}{n + 1}, \quad a + 2\,\frac{b - a}{n + 1}, \quad \text{\&c.}$$

Ex. 4. Find an arithmetical series in which the seventh term is three times as great as the second, and the fourth exceeds the second by four.

If a be the first term, d the common difference, we have the conditions

$$a + 6d = 3(a + d),$$

$$a + 3d = a + d + 4,$$

or $$2a = 3d,$$

$$2d = 4;$$

$$\therefore\ d = 2, \quad a = 3,$$

and the series is 3, 5, 7, 9, &c.

ON GEOMETRICAL PROGRESSION.

103. Def. A series of quantities are said to be in geometrical progression when each term of the series is equal to that which precedes it multiplied by some constant factor, *i.e.* some factor which is the same for all the terms, or in other words, when the ratio of any two successive terms is the same.

Thus a, ar, ar^2, ar^3, ... is a geometrical series.

104. *To sum a geometrical series.*

Let a be the first term, r the common ratio of the terms; then the second term will be ar, the third ar^2, and generally the n^{th} term will be ar^{n-1}.

Let S be the sum of n terms, then

$$S = a + ar + ar^2 + \ldots\ldots\ldots + ar^{n-1},$$

multiplying by r we have

$$rS = ar + ar^2 + \ldots\ldots\ldots + ar^{n-1} + ar^n,$$

subtracting the former of these equations from the latter, we have

$$(r - 1)\,S = ar^n - a;$$

$$\therefore\ S = a\,\frac{r^n - 1}{r - 1},$$

which is the expression for the sum required.

COR. Any three of the four quantities a, r, n, S being given, the fourth may be found.

105. If the quantity r be less than 1, r^n will be less as n is greater, and if we suppose n *indefinitely* great, r^n will be indefinitely small, and we shall have

$$S = \frac{a}{1 - r}.$$

Hence it appears that if we have a geometrical series in which the common ratio is a proper fraction there is a certain quantity, viz. $\frac{a}{1-r}$, to which the sum of the series continually approximates the more terms we take, and consequently we may say that the sum of such a series extended *ad infinitum* is $\frac{a}{1-r}$.

106. An example of a geometrical series continued *ad infinitum* occurs in arithmetic, in the case of recurring decimals: thus the recurring decimal $.\dot{3}33\ldots\ldots = \frac{3}{10} + \frac{3}{10^2} + \frac{3}{10^3} + \ldots\ldots,$

and the sum of this series is $\dfrac{\frac{3}{10}}{1 - \frac{1}{10}} = \frac{3}{9} = \frac{1}{3}$.

But more generally,

107. *To find the vulgar fraction corresponding to given recurring decimal.*

Let the decimal be represented by $A\,.\,BRRR\,\ldots\ldots\ldots$ where A is the integral part, B the nonrecurring decimal part, and R the recurring; and suppose B to contain p digits, and R to contain q digits.

$$\text{Let } S = A\,.\,BRRR\,\ldots\ldots\ldots;$$

$$\therefore\ 10^{p+q}S = ABR\,.\,RR\ldots\ldots\ldots,$$

$$\text{and } 10^{p}S = AB\,.\,RR\ldots\ldots\ldots;$$

$$\therefore\ (10^{p+q} - 10^{p})\,S = ABR - AB,$$

$$\text{and } S = \frac{ABR - AB}{10^{p}(10^{q}-1)}.$$

Ex. 1. Find the sum of 10 terms of the geometrical series 1, 2, 4, 8, ...

In this case $a = 1$, $r = 2$, $n = 10$;

$$\therefore\ S = \frac{2^{10} - 1}{2 - 1} = 2^{10} - 1 = 1023.$$

Ex. 2. There is a geometrical series of which the *second* term is 6 and the *fourth* 54; find it.

Here
$$ar = 6,$$
$$ar^3 = 54;$$
$$\therefore\ r^2 = \frac{54}{6} = 9,$$
$$r = \pm 3;$$
$$\therefore\ a = \frac{6}{\pm 3} = \pm 2;$$

$\therefore$ the series is

$$2,\ 6,\ 18,\ 54,\ldots\ldots\ldots$$
$$\text{or } -2,\ 6,\ -18,\ 54,\ldots\ldots\ldots$$

Ex. 3. Insert n geometrical means between a and b. This is in other words to construct a series of which the first term is a, and the $(n + 2)^{\text{th}}$ term b.

Therefore we must have

$$ar^{n+1} = b,$$

$$\therefore\ r = \left(\frac{b}{a}\right)^{\frac{1}{n+1}},$$

and the geometrical means required are

$$a\left(\frac{b}{a}\right)^{\frac{1}{n+1}},\quad a\left(\frac{b}{a}\right)^{\frac{2}{n+1}},\quad \ldots\ldots a\left(\frac{b}{a}\right)^{\frac{n}{n+1}}.$$

Ex. 4. Find the sum of the series $1 + \frac{1}{2} + \frac{1}{4} + \ldots$ *ad infinitum.*

In this case $\qquad a = 1,\quad r = \frac{1}{2};$

$$\therefore\ S = \frac{1}{1 - \frac{1}{2}} = 2.$$

Ex. 5. Find the vulgar fraction corresponding to the recurring decimal $2.46\dot{2}\dot{6}2\ldots\ldots$

Let $\qquad S = 2.46\overline{262},$

$$1000S = 2462.\overline{62},$$

$$10S = \quad 24.\overline{62};$$

$$\therefore\ 990S = 2438,$$

$$\text{and } S = \frac{2438}{990} = \frac{1219}{495}.$$

ON HARMONICAL PROGRESSION.

108. Def. A series of quantities are said to be in harmonical progression, when any three successive terms are so related, that the first is to the third as the difference between the first and the second is to the difference between the second and third.

Thus if a, b, c are in harmonical progression,

$$a : c :: a - b : b - c.$$

The reciprocals of quantities in harmonical progression are in arithmetical progression.

Let a, b, c be quantities in harmonical progression, then by definition

$$\frac{a}{c} = \frac{a-b}{b-c},$$

$$\text{or } \frac{b-c}{c} = \frac{a-b}{a},$$

$$\text{or } \frac{1}{c} - \frac{1}{b} = \frac{1}{b} - \frac{1}{a},$$

which proves that the difference between $\frac{1}{a}$ and $\frac{1}{b}$ is the same as between $\frac{1}{b}$ and $\frac{1}{c}$, or that $\frac{1}{a}$, $\frac{1}{b}$, $\frac{1}{c}$ are in arithmetical progression.

COR. Hence we may, if we please, take it as the definition of quantities in harmonical progression that their reciprocals are in arithmetical.

109. A series of quantities in harmonical progression admits of no simple summation.

110. The three kinds of progression which have been treated of, may be brought under one point of view as follows:

If a, b, c are in *arithmetical* progression, we have

$$\frac{a-b}{b-c} = \frac{a}{a}.$$

If in *geometrical*,

$$\frac{a-b}{b-c} = \frac{a}{b}.$$

If in *harmonical*,

$$\frac{a-b}{b-c} = \frac{a}{c}.$$

ON PERMUTATIONS AND COMBINATIONS.

111. The different ways in which any number of quantities can be arranged are called their *permutations*.

Thus the *permutations* of the letters a, b, c taken two together are ab, ac, ba, bc, ca, cb.

The *combinations* of a number of quantities are the collections which can be made of them without regard to arrangement.

Thus the *combinations* of a, b, c, taken two together, are ab, ac, bc : ab, ba, which were *two permutations*, form only *one combination*, and so of the rest.

112. *To find the number of permutations of* n *things taken* r *together.*

Let $\{nPr\}$ denote the number of permutations of n things a, b, c, d...... taken r together. Now suppose we omit one of the letters, as a, and form the remainder into permutations taken $r-1$ together, of which, according to our notation, there will be $\{(n-1)P(r-1)\}$; then before each of these permutations we may place a, *thus forming* $\{(n-1)P(r-1)\}$ *permutations of n things taken r together, in which a stands first:* the same may be said of b, c, d, and there are n of them; therefore we shall have

$$\{nPr\} = n\,\{(n-1)P(r-1)\};$$

in like manner,

$$\{(n-1)P(r-1)\} = (n-1)\,\{(n-2)P(r-2)\},$$

$$\{(n-2)P(r-2)\} = (n-2)\,\{(n-3)P(r-3)\},$$

$$\&c. = \&c.$$

$$\begin{aligned}\{(n-r+2)P2\} &= (n-r+2)\,\{(n-r+1)P1\}\\ &= (n-r+2)(n-r+1),\end{aligned}$$

[since it is manifest that $\{(n-r+1)P1\} = n-r+1$.]

Now multiplying together the corresponding sides of these equations and leaving out the common factors, we have

$$\{nPr\} = n(n-1)(n-2)\ldots\ldots(n-r+2)(n-r+1).$$

Cor. If $r = n$, we have

$$\{nPn\} = n(n-1)(n-2)\ldots\ldots 2.1.$$

113. *To find the number of permutations of* n *things taken all together, when* a *are of the same kind.*

Let x be the number required.

Then since *all* the a quantities enter into *each* permutation, if we suppose them all different, *each* permutation would be resolved into $\{aPa\}$ permutations, and therefore the whole number of permutations would be $\{aPa\}$ times as great; but in this case the number of permutations would be that of n things, all different, taken all together, or $\{nPn\}$; hence we have

$$x\{aPa\} = \{nPn\},$$

$$\text{or } x = \frac{n(n-1)\ldots\ldots 2.1}{a(a-1)\ldots\ldots 2.1}.$$

Cor. In like manner, if there were a quantities of one kind, β of another, γ of another, &c. we should have for the number of permutations

$$\frac{n(n-1)\ldots\ldots 2.1}{1.2\ldots a.1.2\ldots\beta.1.2\ldots\gamma.\&c.}.$$

114. *To find the number of combinations of* n *things taken* r *together.*

Let $\{nCr\}$ denote the number of combinations. Then since the order of the quantities is not regarded in a combination, each combination of r quantities may be resolved by permuting them into $\{rPr\}$ permutations. Hence we have

$$\{nCr\}\{rPr\} = \{nPr\},$$

$$\text{or } \{nCr\} = \frac{n(n-1)\ldots\ldots(n-r+1)}{1.2\ldots\ldots\ldots\ldots r}.$$

115. *The number of combinations of* n *things taken* r *together is the same as that of* n *things taken* (n – r) *together.*

In other words, $\{nCr\} = \{nC(n-r)\}$.

$$\text{Now } \{nCr\} = \frac{n(n-1)\ldots\ldots(n-r+1)}{1.2\ldots\ldots\ldots\ldots\ldots r},$$

$$\{nC(n-r)\} = \frac{n(n-1)\ldots\ldots(r+1)}{1.2\ldots\ldots\ldots\ldots(n-r)}.$$

Suppose $n-r$ greater than r, which may be done without affecting the generality of the proof:

$$\text{then } \{nC(n-r)\} = \frac{n(n-1)\ldots\ldots(r+1)}{1.2\ldots\ldots\ldots\ldots n-r}$$

$$= \frac{n(n-1)\ldots(n-r+1)(n-r)\ldots\ldots(r+1)}{1.2\ldots\ldots r(r+1)\ldots\ldots\ldots\ldots\ldots(n-r)}$$

$$= \frac{n(n-1)\ldots\ldots\ldots(n-r+1)}{1.2\ldots\ldots\ldots\ldots\ldots r} = (nCr).$$

116. Another mode of proving this proposition is as follows: whenever r of the n quantities are taken to form a combination, $n-r$ are necessarily omitted, and these may be supposed to be formed into a combination which may be called with reference to the former a *complementary* combination. Hence each combination of r quantities has its complementary combination of $n-r$, and therefore the number of the two sets of combinations is equal.

ON THE BINOMIAL THEOREM.

117. We have already seen (Art. 46.) that $(a+b)^2 = a^2 + 2ab + b^2$, and that $(a+b)^3 = a^3 + 3a^2b + 3ab^2 + b^3$, and we might find the expansion of any other positive integral power of $a+b$ by actual multiplication: the binomial theorem

is a formula for the general expansion of $(a+b)^n$ according to powers of b, and that not only in the case of n being positive and integral, but also when it is fractional and negative.

118. *To investigate the Binomial Theorem in the case of a positive integral index.*

We have by actual multiplication

$$(x+a)(x+b) = x^2 + (a+b)x + ab,$$

$$(x+a)(x+b)(x+c) = x^3 + (a+b+c)x^2 + (ab+ac+bc)x + abc.$$

In examining the preceding expressions we observe the following laws:

(1) That they consist of a series of descending powers of x, and that the first and highest index is the number of factors forming the expression.

(2) That the coefficient of the first term is unity; of the second, the sum of the products of the quantities a, b, c, taken *one* together; of the third, the sum of the products of the same taken *two* together; of the last, the product of them taken all together.

Let us suppose that this law holds for n factors, that is, that

$$(x+a)(x+b)(x+c)\dots(x+p) = x^n + S_1x^{n-1} + S_2x^{n-2} + \dots + S_n,$$

$$\text{where } S_1 = a+b+c+\dots+p,$$

$$S_2 = ab + ac + \dots$$

$$\vdots$$

$$S_n = abc\dots p.$$

Then will

$$(x+a)(x+b)(x+c)\dots(x+p)(x+q)$$

$$= (x+q)\left\{x^n + S_1x^{n-1} + S_2x^{n-2} + \dots + S_n\right\}$$

$$= x^{n+1} + S_1x^n + S_2x^{n-1} + \dots + S_nx$$

$$+ qx^n + qS_1x^{n-1} + \dots + qS_n$$

$$= x^{n+1} + S_1'x^n + S_2'x^{n-1} + \dots + S_n' \text{ suppose,}$$

where $S_1' = S_1 + q = a + b + c + \ldots + p + q$,

$$S_2' = S_2 + qS_1 = ab + ac + \ldots + qa + qb + \ldots$$

$$S_n' = qS_n = abc\ldots pq.$$

Hence it appears, that if the assumed law be true for n factors, it will be true for $n+1$; but it *is* true for *three*; $\therefore$ for *four*; $\therefore$ &c. $\therefore$ generally true.

Now let $a = b = c = \ldots = p$.

Then $S_1 = a + a + a + \ldots$ to n terms $= na$,

$S_2 = a^2 + a^2 + \ldots$ to as many terms as there are combinations of n things taken *two* together

$$= \frac{n(n-1)}{1.2} a^2, \text{ (Art. 114.)}$$

&c. = &c.

$$S_n = a^n;$$

and $(x+a)(x+b)(x+c)\ldots(x+p)$ becomes $= (x+a)^n$;

$$\therefore (x+a)^n = x^n + nx^{n-1}a + \frac{n(n-1)}{1.2} x^{n-2}a^2 + \ldots + a^n.$$

The general term being $\dfrac{n(n-1)\ldots(n-r+1)}{1.2\ldots\ldots\ldots\ldots r} x^{n-r}a^r$, and the number of terms $n+1$.

Cor. 1. $(1+x)^n = 1 + nx + \dfrac{n(n-1)}{1.2} x^2 + \ldots + x^n$.

Cor. 2. It appears from the proposition proved in Art. 115. that the coefficients of terms at the same distance from the beginning and end of the series are the same; which is also otherwise apparent from the fact that $(x+a)^n = (a+x)^n$.

Ex. 1. $(a+b)^4 = a^4 + 4a^3b + \dfrac{4.3}{1.2} a^2b^2 + \dfrac{4.3.2}{1.2.3} ab^3 + b^4$,

$$= a^4 + 4a^3b + 6a^2b^2 + 4ab^3 + b^4.$$

Ex. 2. Find the coefficient of $a^5 b^3$ in the expansion of $(a+b)^8$.

$$\text{The coefficient} = \frac{8 \cdot 7 \cdot 6}{1 \cdot 2 \cdot 3} = 56.$$

119. The series which has been proved for $(1+x)^n$ when n is positive and integral may be shewn to be the true series in the case of n being fractional or negative; it is to be remarked, however, that there is this important distinction between the two cases, that when n is a positive integer the series comes to an end, but when n is fractional or negative the series will be infinitely extended. In fact we have seen that the general term of the series, when n is a positive integer, is

$$\frac{n \cdot (n-1) \ldots\ldots\ldots (n-r+1)}{1 \cdot 2 \ldots\ldots\ldots r},$$

and this becomes zero when

$$r = n + 1;$$

but if we prove that the same is the general term when n is fractional or negative, it will be apparent that it never can become zero, and therefore the series can never terminate. We may, however, consider the series indefinitely continued even in the case of a being a positive integer, only that after a certain number the terms will be in reality evanescent.

This being premised, we may proceed

120. *To extend the Binomial Theorem to the case of fractional and negative indices.*

Let the series

$$1 + mx + \frac{m(m-1)}{1 \cdot 2} x^2 + \&c.,$$

(*where m may be any quantity whatever*) be represented for shortness' sake by the symbol $f(m)$.

Then, according to the same notation,

$$f(n) = 1 + nx + \frac{n(n-1)}{1 \cdot 2} x^2 + \ldots\ldots\ldots$$

Our first step will be to determine the form of the product $f(m) \times f(n)$: to do this we observe, that by actual multiplication it is clear that $f(m) \times f(n)$ will be a series proceeding by ascending powers of x,

$$= 1 + Ax + Bx^2 + \dots\dots\dots \text{ suppose.}$$

It would be possible to determine the coefficients A, B... by actual multiplication; but we obtain them more simply by this consideration, that although the *values* of A, B, ... are altered by altering m and n, yet their *forms*, that is, the manner in which m and n are involved in them, are the same whatever m and n may be; and therefore, if we discover the form of the product $f(m) \times f(n)$ in the case of m and n being positive integers, we shall know its form whatever m and n may be.

$$\text{But in that case} \quad f(m) = (1 + x)^m,$$

$$\text{and } f(n) = (1 + x)^n;$$

$$\therefore f(m) \times f(n) = (1 + x)^{m+n},$$

$$= f(m + n) \text{ by the notation;}$$

and hence, by the preceding reasoning, we must have *universally*

$$f(m) \times f(n) = f(m + n);$$

and in like manner we shall have for any number of factors,

$$f(m) \times f(n) \times f(p) \times \dots\dots\dots = f(m + n + p + \dots\dots\dots)$$

This being premised, in the formula

$$f(m) \times f(n) \times f(p) \dots\dots\dots = f(m + n + p + \dots\dots\dots),$$

$$\text{make } m = n = p = \dots\dots\dots = \frac{\mu}{\nu},$$

where μ and ν are positive whole numbers, and let there be ν factors; then we have

$$\left\{f\left(\frac{\mu}{\nu}\right)\right\}^\nu = f(\mu)$$

$$= (1 + x)^\mu \text{ since } \mu \text{ is a positive integer;}$$

$$\therefore\ (1+x)^{\frac{\mu}{\nu}} = f\left(\frac{\mu}{\nu}\right) = 1 + \frac{\mu}{\nu}x + \frac{\frac{\mu}{\nu}\left(\frac{\mu}{\nu}-1\right)}{1\,.\,2}x^2 + \ldots\ldots$$

which proves the theorem for fractional indices.

Again, in the formula

$$f(m) \times f(n) = f(m+n),$$

let $m = -n$, and let n be positive;

$$\therefore\ f(-n) \times f(n) = f(0) = 1;$$

$$\therefore\ f(-n) = \frac{1}{f(n)} = \frac{1}{(1+x)^n} = (1+x)^{-n};$$

$$\therefore\ (1+x)^{-n} = f(-n) = 1 - nx + \frac{-n(-n-1)}{1\,.\,2}x^2 - \&c.$$

which proves the theorem for negative indices.

Ex. 1.

$$(1+x)^{-2} = 1 - 2x + \frac{-2(-2-1)}{1\,.\,2}x^2 + \frac{-2(-2-1)(-2-2)}{1\,.\,2\,.\,3}x^3 + \&c.$$

$$= 1 - 2x + \frac{2\,.\,3}{1\,.\,2}x^2 - \frac{2\,.\,3\,.\,4}{1\,.\,2\,.\,3}x^3 + \&c.$$

$$= 1 - 2x + 3x^2 - 4x^3 + \&c.$$

Ex. 2. $\qquad (a+bx)^{-1} = \dfrac{1}{a}\left(1 + \dfrac{bx}{a}\right)^{-1}$

$$= \frac{1}{a}\left\{1 - \frac{bx}{a} + \frac{-1(-1-1)}{1\,.\,2}\frac{b^2x^2}{a^2} + \frac{-1(-1-1)(-1-2)}{1\,.\,2\,.\,3}\frac{b^2x^3}{a^3} + \ldots\right\}$$

$$= \frac{1}{a}\left\{1 - \frac{bx}{a} + \frac{1\,.\,2}{1\,.\,2}\frac{b^2x^2}{a^2} - \frac{1\,.\,2\,.\,3}{1\,.\,2\,.\,3}\frac{b^3x^3}{a^3} + \ldots\right\}$$

$$= \frac{1}{a}\left(1 - \frac{bx}{a} + \frac{b^2x^2}{a^2} - \frac{b^3x^3}{a^3} + \ldots\ldots\right),$$

which is a result obtainable by actual division; it may be observed, however, that if the result had been so obtained we should have had also a *remainder*, of which there is no trace in the preceding series: the fact is, that the series obtained by the Binomial Theorem can only be considered as numerically equal to the quantities expanded, when the series are *convergent*. When the convergence is very rapid the Binomial Theorem may be conveniently used for purposes of approximation, as will be seen in the next Example.

Ex. 3. To find an approximate value of $\sqrt[p]{N^p + x}$, when x is much smaller than N^p.

$$(N^p + x)^{\frac{1}{p}} = N\left(1 + \frac{x}{N^p}\right)^{\frac{1}{p}}$$

$$= N\left\{1 + \frac{1}{p}\,\frac{x}{N^p} + \frac{\frac{1}{p}\left(\frac{1}{p} - 1\right)}{1.2}\,\frac{x^2}{N^{2p}} + \ldots\ldots\right\}.$$

By taking a few terms of this series we may obtain the result with a considerable degree of accuracy.

Ex. 4. To find an approximate value of $\sqrt{50}$.

$$\sqrt{50} = (49 + 1)^{\frac{1}{2}} = 7\left(1 + \frac{1}{49}\right)^{\frac{1}{2}}$$

$$= 7\left\{1 + \tfrac{1}{2}\,\frac{1}{49} + \frac{\frac{1}{2}\left(\frac{1}{2} - 1\right)}{1.2}\,\frac{1}{49^2} + \frac{\frac{1}{2}\left(\frac{1}{2} - 1\right)\left(\frac{1}{2} - 2\right)}{1.2.3}\,\frac{1}{49^3} + \ldots\right\}$$

$$= 7\left(1 + \frac{1}{2}\,\frac{1}{49} - \frac{1}{8}\,\frac{1}{49^2} + \frac{1}{16}\,\frac{1}{49^3} - \ldots\right)$$

$$= 7\left(1 + \frac{.020408}{2} - \frac{.000416}{8} + \frac{.000008}{16} - \ldots\right)$$

$$= 7\,(1.010204 - .000052 + \ldots)$$

$$= 7\,(1.010152) = 7.071064.$$

which is correct to five places of decimals.

ON LOGARITHMS.

121. Def. The logarithm of a number N is the value of x which satisfies the equation $a^x = N$, where a is some given number.

Thus if a be 10, the logarithm of 100 is 2, that of 1000 is 3; and that of any number between 100 and 1000 will be greater than 2 and less than 3, so that it may be represented by 2 followed by places of decimals.

If we give a any value, as 10, it is possible to find the values of x corresponding to all values of N, that is, to find the logarithm of all numbers to the *base* 10; suppose these found and registered in tables, these will be the common tables of logarithms; we shall see of what use they may be made from the following propositions.

122. *The logarithm of the product of two numbers is the sum, and the logarithm of the quotient is the difference of the logarithms of the numbers.*

For let $a^x = N$, $a^y = N'$, where N N' are any two numbers, and x y their logarithms to the base a,

$$\text{then } a^{x+y} = NN',$$

but by definition $x + y$ is the logarithm of NN' to base a, or (as we usually write it) $x + y = \log_a NN'$;

$$\therefore \log_a NN' = \log_a N + \log_a N'.$$

In like manner

$$\frac{N}{N'} = a^{x-y},$$

$$\text{and } \therefore \log_a \frac{N}{N'} = \log_a N - \log_a N'.$$

123. *The logarithm of any power of a number is the logarithm of the number multiplied by the index which expresses the power.*

Suppose $a^x = N,$

$$\text{then } a^{px} = N^p,$$

$$\text{or } px = \log_a N^p \text{ by definition,}$$

$$\text{but } x = \log_a N;$$

$$\therefore \log_a N^p = p \log_a N.$$

In like manner

$$\log_a N^{\frac{1}{p}} = \frac{1}{p} \log_a N.$$

124. Hence it appears that by means of a table of logarithms, *multiplication* may be performed by addition, *division* by subtraction, *involution* by multiplication, and *evolution* by division.

For suppose that we possess such a table, and that we wish to multiply together two numbers N and N'. We look for the logarithms of these two numbers, add them together, and then look among the logarithms for the sum thus found, the number corresponding to that logarithm will be NN': and so of the other operations. From this it will be easily understood that the use of logarithms greatly facilitates long calculations.

125. *To explain the advantage of choosing* 10 *as the base of a system of logarithms.*

Suppose we have any two numbers in which the digits are the same, but which differ from each other in the position of the unit's place: for example, 137 and 13700. Then

$$13700 = 137 \times 100 = 137 \times 10^2;$$

$$\therefore \log_{10} 13700 = \log_{10} 137 + 2.$$

Hence the logarithms of the two numbers in question differ

from each other only in this, that the larger one has 2 added to it, the decimal parts of the two being the same. And we may, in like manner, conclude that the decimal parts of the logarithms of all numbers having the same digits, but a different unit's place, are the same. Hence, if we have a rule for assigning the integral part of a logarithm, the tables need not contain the logarithms of all numbers, but only of those in which the digits are different.

The integral part of a logarithm is called the *characteristic*, the decimal part the *mantissa*.

126. *To find a rule for ascertaining the characteristic of the logarithm of any number.*

We have $\log_{10} 10^2 = 2$, and $\log_{10} 10^3 = 3$; hence the logarithm of any number between 100 and 1000, i.e. of any number composed of three digits, is between 2 and 3, and therefore the characteristic is 2; similarly for numbers of four digits the characteristic is 3; and generally the characteristic is one less than the number of digits. If the number be decimal we must have a negative characteristic, for

$$\log_{10} 1 = 0$$

$$\log_{10} \tfrac{1}{10} = \log_{10} .1 = -1,$$

$$\text{and } \log_{10} \tfrac{1}{100} = \log_{10} .01 = -2;$$

hence the logarithm of a number between 1 and .1 is less than 0 and greater than -1, and may therefore be represented by $-1 +$ a mantissa; in like manner the logarithm of a number between .1 and .01 will be $-2 +$ a mantissa; and generally, the characteristic will be one greater than the number of cyphers which precede the first significant figure.

127. The preceding articles contain the principal properties of logarithms, which constitute their utility; the actual calculation of them would involve us in series with which the student is not at present acquainted, and for which, if he be desirous of pursuing the subject, he is referred to other treatises.

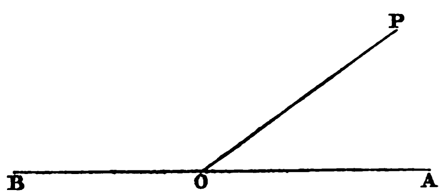
PLANE TRIGONOMETRY.

6

PLANE TRIGONOMETRY.

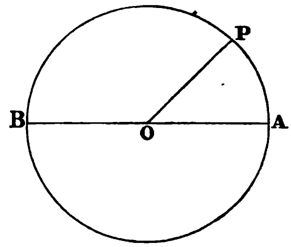
1. The science of *plane trigonometry*, according to the strict meaning of the words, treats of the *measurement* of *plane triangles*; we may however consider the name as applicable to the more general subject of the measurement of plane *angles*, of which the measurement of triangles forms an important part.

2. The term *angle* will be used in this subject in a more extended sense than that which is attached to it in Euclid's elements, for an angle according to Euclid's definition cannot exceed two right angles, and indeed, according to our ordinary conception of an *angle* or *corner*, it is manifest that there can be no such thing as an angle exceeding that limit: but there is no such restriction in Trigonometry; in that science the magnitude of an angle is unlimited. To make this understood, let BOA be a fixed straight line, and OP a line which revolves about O, and which at first coincided with OA. Then we say, that when OP is in the position represented in the figure, it has described the angle AOP; but this mode of conceiving an angle admits of extension to angles of any magnitude, for we may suppose OP to revolve beyond OB and so describe an angle greater than two right angles, or more generally we may suppose it to describe an angle of any magnitude whatever.



3. The same thing may be put in a slightly different point of view, by considering the point P to trace out a circle with centre O. Then it is proved by Euclid, (VI. 33,) that in the same circle the angle standing on any arc is proportional

to that arc, so that the arc subtending an angle is a proper *measure* of that angle. Now it will be seen that the arc traced out by P may be perfectly unlimited, may in fact consist of any number of circumferences or any portion of a circumference. The student will frequently find it convenient to have before his mind the notion of the subtending *arc*, as the *measure* of the subtended angle.



ON THE MODE OF MEASURING ANGLES.

4. Suppose a right angle to be divided into ninety equal parts, and let each part be called a *degree*; suppose a *degree* to be divided into *sixty* parts called *minutes*, and a minute into *sixty* parts called *seconds*: then the magnitude of an angle may be assigned by saying how many degrees, minutes, and seconds it contains.

Degrees, minutes, and seconds, are thus denoted, $25^0\ 30'\ 32''$. We may, if we please, denote parts of a second by a similar notation; but in general an angle will be determined with sufficient accuracy, by assigning the degrees, minutes, and seconds, which it contains, or we may make the value exact by setting down besides fractions of a second.

According to this notation, 90^0 represents a right angle, 180^0 two right angles, and 360^0 four right angles or the angle described by a complete revolution of the revolving line.

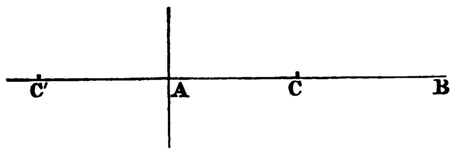
5. It is obvious, that this mode of measuring angles is perfectly arbitrary; we might, for instance, (as some French authors have done) divide the right angle into one hundred degrees, each degree into one hundred parts, and so on; and this is, in fact, a very convenient division, because no new notation will be required to denote the different parts; for such a quantity as $3^0\ 45'\ 17''$ might be written 3.4517, and thus the

only mark required is a decimal point to separate the degrees from the parts. Nevertheless as the former division is universal in this country, we shall adopt it here. Another mode of measuring angles will be given hereafter. (See Art. 52, page 123.)

ON THE USE OF THE SIGNS + AND − TO INDICATE THE DIRECTIONS OF LINES.

6. The primary use of the signs + and − is, as we have seen, (Algebra, Art. 5.) to denote addition and subtraction; nevertheless, we found that these signs immediately introduced the notion of *negative* quantities, and we illustrated the meaning of a negative quantity by a debt, which may be looked upon as a quantity *to be* subtracted.

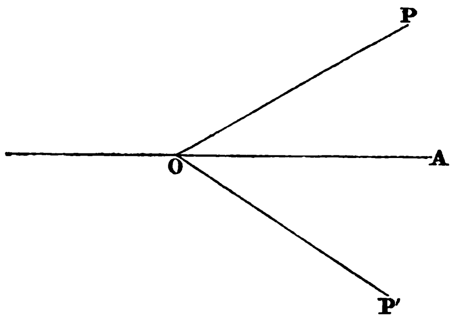
We must now still further generalize the meaning of the signs + and −, and it will be easy to shew that if a line drawn in one direction be called *positive*, then a line drawn in the



opposite direction will properly be accounted *negative*. To make this appear, let AB be any straight line of length a, and let us cut off from it a portion $CB = b$, then the remainder $AC = a - b$. Now so long as b is less than a, C lies to the right of A and $a - b$ is positive; but suppose we endeavour to cut off from AB a portion BC' greater than itself, then (although the operation is really impossible, yet) according to the analogy of the preceding operation we shall have for the remainder AC', which is a line drawn to the *left* of A and is represented by $-(b - a)$, or by a *negative* quantity. Hence then we see how that if a line drawn in one direction is accounted positive, then a line drawn in the opposite direction is properly accounted negative.

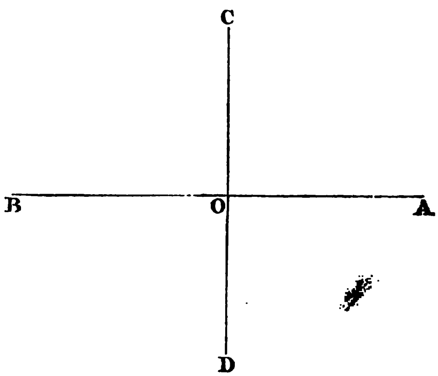
7. A more comprehensive view of this subject may be given, by saying that the signs + and − denote any exactly opposite qualities of a quantity; we are at liberty to take these signs in such a sense, because it includes the original meaning, viz. that of addition and subtraction, for *additive* and *subtractive* are qualities exactly the reverse of each other. And if we take the signs + and − as having this meaning, we shall see at once, that among other qualities, they properly designate opposition in *direction* when applied to lines.

As an example, let us consider what will be the meaning of a negative *angle*. Suppose the line OP by its revolution about O and *upwards* from OA to describe the angle AOP, and let angles described in this manner be considered *positive*. Then if the line revolve *downwards* from OA and describe the angle AOP', this angle will be properly accounted negative.



The same explanation would apply to negative *arcs*.

8. In what follows, we shall consider that if AOB, COD, are two lines at right angles to each other, then lines drawn parallel to AOB are positive if to the right, negative if to the left of CD, and lines drawn parallel to COD are positive if drawn above AOB, and negative if below it.



We shall see immediately the great advantage of the preceding conventions.

ON THE TRIGONOMETRICAL FUNCTIONS OF AN ANGLE.

Fig. I.

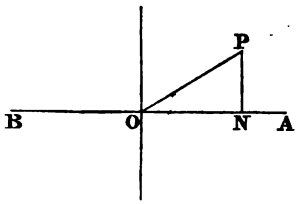


Fig. II.

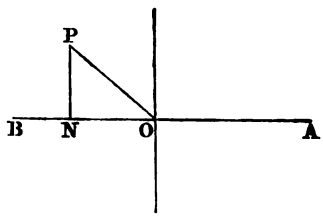


Fig. III.

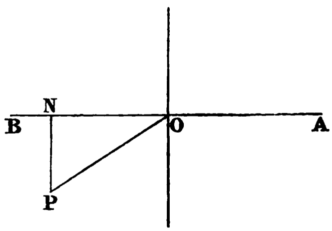
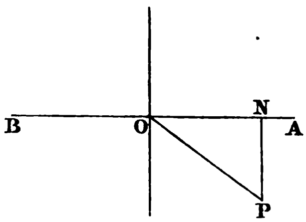


Fig. IV.



9. Let the line OP revolving from the initial position OA about O describe the angle AOP: from P let fall the perpendicular PN upon the line AOB, then the ratio

(1) $\dfrac{PN}{OP}$ is defined to be the *sine* of the angle AOP.

(2) $\dfrac{ON}{OP}$ .... *cosine* ....

(3) $\dfrac{PN}{ON}$ .... *tangent* ....

(4) $\dfrac{ON}{PN}$ .... *cotangent* ....

(5) $\dfrac{OP}{ON}$ .... *secant* ....

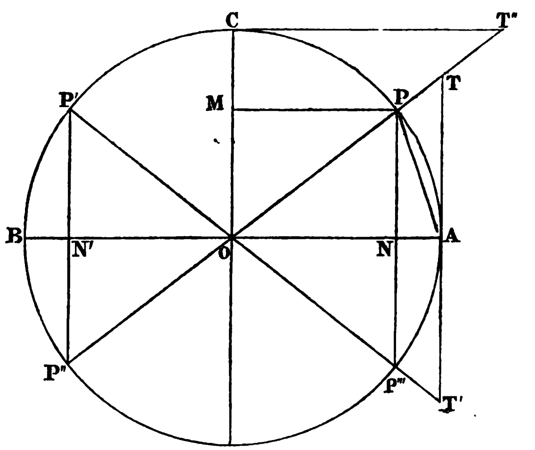
(6) $\dfrac{OP}{PN}$ is defined to be the *cosecant* of the angle AOP.

(7) $1 - \dfrac{ON}{OP}$ .... *versedsine* ....

The ratios to which we have just assigned names are called the trigonometrical *functions* of the angle, that is, quantities which depend upon that angle for their value, and (conversely) which being given determine the value of the angle.

For shortness' sake we usually denote an angle by some single letter, as for instance A; and we write the names of the functions above defined thus, $\sin A$, $\cos A$, $\tan A$, $\cot A$, $\sec A$, $\operatorname{cosec} A$, $\operatorname{vers} A$.

10. The meaning of these names, and the relations of the trigonometrical functions to each other, will be seen more distinctly from another mode of defining them.



Suppose the point P at the extremity of the revolving line OP to trace out a circle of radius r. At the point A, which is the initial position of P, draw the tangent TAT', and let OP be produced to meet AT in T; also draw PN perpen-

dicular to OA; then we may define the trigonometrical function of the angle POA or A, thus:

$$\sin A = \frac{PN}{r},$$

$$\tan A = \frac{AT}{r}\left(= \frac{PN}{ON}, \text{ by similar triangles}\right),$$

$$\sec A = \frac{OT}{r}\left(= \frac{OP}{ON}, \text{ by similar triangles}\right),$$

$$\text{vers} A = \frac{AN}{r}.$$

With regard to the three other functions we may observe that the *complement* of an angle is its defect of a right angle, and that *cosine* merely signifies sine of the *complement*, *cotangent* tangent of the *complement*, and *cosecant* secant of the *complement*. Hence these functions need no new definition; but if we take AOC a right angle and draw the tangent CT'', and PM perpendicular to OC, and produce OP to meet CT'' in T'', we shall have

$$\cos A = \sin COP = \frac{PM}{r},$$

$$\cot A = \tan COP = \frac{CT''}{r} = \frac{PM}{OM} = \frac{ON}{PN},$$

$$\text{cosec} A = \sec COP = \frac{OT''}{r} = \frac{OP}{OM} = \frac{OP}{PN}.$$

Another function is sometimes introduced: if we join AP, then $\frac{AP}{r}$ is called the *chord* of A or chd A.

11. From the definitions given of the trigonometrical functions a number of connecting relations may be easily

deduced; the following are some of the most important, and the student is advised to examine their truth and make himself familiar with them:

$$\sin^2 A + \cos^2 A = 1\text{*},$$

$$\tan A = \frac{\sin A}{\cos A},$$

$$\sec A = \frac{1}{\cos A},$$

$$\cot A = \frac{\cos A}{\sin A},$$

$$\operatorname{cosec} A = \frac{1}{\sin A}.$$

12. Furthermore, if one of the trigonometrical functions be given, it is not difficult to express all the rest in terms of it. For example, let it be required to express all the functions in terms of the *sine*.

We have
$$\cos A = \pm\sqrt{1 - \sin^2 A},$$

$$\tan A = \pm\frac{\sin A}{\sqrt{1 - \sin^2 A}},$$

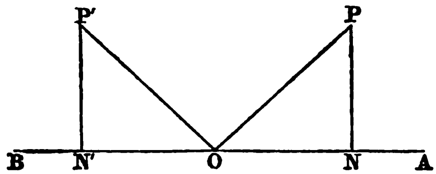
$$\sec A = \pm\frac{1}{\sqrt{1 - \sin^2 A}},$$

$$\cot A = \pm\frac{\sqrt{1 - \sin^2 A}}{\sin A},$$

$$\operatorname{cosec} A = \pm\frac{1}{\sin A}.$$

* The notation $\sin^2 A$ $\cos^2 A$ &c. represents $(\sin A)^2$ $(\cos A)^2$ &c. or the *square* of $\sin A$ $\cos A$ &c. It appears on the whole to be the most convenient notation, though not universally adopted.

13. Let us here inquire by the way what is the meaning of the ambiguity of the *sign* in the preceding example.



The reason will appear thus: let AOP be the angle of which the sine is given; take $BOP' = AOP$, then it is evident that $\sin AOP' = \sin AOP$, and therefore the given sine may as well belong to AOP' as to AOP. But it is not true that $\cos AOP' = \cos AOP$; for ON, ON' are drawn on opposite sides of O, and therefore if one is positive the other is negative: hence we have $\cos AOP' = -\cos AOP$; and therefore in the above set of formulæ, we must take the upper or lower sign, according as we suppose the given sine to belong to AOP or AOP'.

14. The difference between a given angle and two right angles is called its *supplement*. The preceding remarks shew us that the sine of an angle is the sine of its supplement, and that the cosine of an angle is the cosine of its supplement with the sign changed.

15. *To trace the signs of sin* A, *cos* A, *tan* A, *sec* A, *as* A *increases from* 0^0 *to* 360^0. (See the figures of Art. 9.)

By our definition, $\sin A = \dfrac{PN}{OP}$, and has therefore the same sign as PN, since there is no reason why OP should ever change sign.

Hence $\sin A$ is positive when A is between 0^0 and 180^0, negative when between 180^0 and 360^0.

Cos $A = \dfrac{ON}{OP}$, and has therefore the same sign as ON.

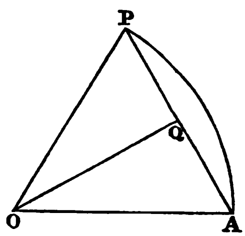
Hence $\cos A$ is positive when A is between 0^0 and 90^0, negative when between 90^0 and 270^0, and positive when between 270^0 and 360^0.

Tan $A = \frac{PN}{ON}$, and is therefore positive when PN and ON have the same sign, negative when they have contrary signs.

Hence tan A is positive when A is between 0 and 90^0, or between 180^0 and 270^0; negative when it is between 90^0 and 180^0, or between 270^0 and 360^0.

Sec $A = \frac{OP}{ON}$, and has therefore the same sign as cos A.

In like manner may be determined the signs of cot A and cosec A. Vers A is always positive.



The sign of chd A may be determined thus. Let AOP be the angle A, AP the subtending arc; join AP, then chd $A = \frac{AP}{AO}$. Now draw OQ perpendicular to AP, and therefore bisecting it; then OQ also bisects the angle AOP, and therefore each of the angles AOQ, $POQ = \frac{A}{2}$;

$$\therefore \text{ chd } A = \frac{AP}{AO} = 2\frac{AQ}{AO} = 2 \sin \frac{A}{2}.$$

Hence chd A has the same sign as $\sin \frac{A}{2}$: but $\sin \frac{A}{2}$ is positive while $\frac{A}{2}$ is between 0^0 and 180^0, or A between 0^0 and 360^0, and negative while $\frac{A}{2}$ is between 180^0 and 360^0, or A between 360^0 and 720^0. Therefore chd A is positive while the revolving line makes its first revolution, negative while it makes its second, and so on.

16. *To determine the change of magnitude of sin* A, *cos* A, *tan* A, *sec* A, *cot* A, *cosec* A, *while* A *increases from* 0^0 *to* 360^0.

Retaining the same figures as in the last article, it will be seen that as

A increases from 0^0 to 90^0, ON decreases from OP to 0,
PN increases from 0 to OP,
A.............. 90 to 180^0, ON increases from 0 to OP,
PN decreases from OP to 0,
A.............. 180^0 to 270^0, ON decreases from OP to 0,
PN increases from 0 to OP,
A.............. 270^0 to 360^0, ON increases from 0 to OP,
PN decreases from OP to 0.

Hence observing the changes of *sign*, as already explained, it is easy to deduce the following table for the changes of sign and value of the functions.

A between	$0 \ldots 90^0$	$90^0 \ldots 180^0$	$180^0 \ldots 270^0$	$270^0 \ldots 360^0$
$\sin A$	$0 \ldots 1$	$1 \ldots 0$	$0 \ldots -1$	$-1 \ldots 0$
$\cos A$	$1 \ldots 0$	$0 \ldots -1$	$-1 \ldots 0$	$0 \ldots 1$
$\tan A$	$0 \ldots +\infty$	$-\infty \ldots 0$	$0 \ldots +\infty$	$-\infty \ldots 0$
$\sec A$	$1 \ldots +\infty$	$-\infty \ldots -1$	$-1 \ldots -\infty$	$+\infty \ldots 1$
$\cot A$	$+\infty \ldots 0$	$0 \ldots -\infty$	$+\infty \ldots 0$	$0 \ldots -\infty$
$\operatorname{cosec} A$	$+\infty \ldots 1$	$1 \ldots +\infty$	$-\infty \ldots -1$	$-1 \ldots -\infty$

The student is recommended to make himself as familiar as possible with the results of the preceding table: and it may be observed to aid him in doing so, that if he becomes well acquainted with the changes of the *sine* and *cosine*, those of the other functions are at once deducible.

It will be remarked that the trigonometrical functions change sign in passing through 0 and ∞, and for no other values.

17. The values of vers A and chd A have not yet been given. They are of no great importance, as those functions may always be replaced by $1 - \cos A$, and $2 \sin \frac{A}{2}$ respectively; they are however as follows:

A between	$0^0 \ldots 90^0$	$90^0 \ldots 180^0$	$180^0 \ldots 270^0$	$270^0 \ldots 360^0$
vers A	$0 \ldots 1$	$1 \ldots 2$	$2 \ldots 1$	$1 \ldots 0$

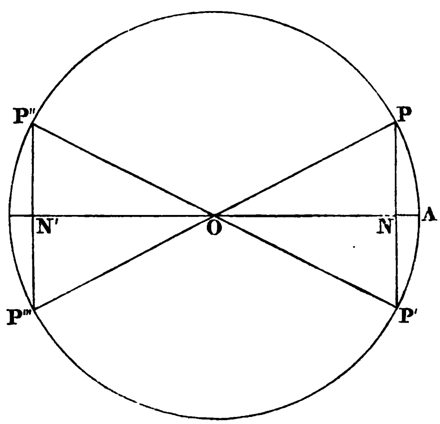
A between	$0^0 \ldots 180^0$	$180^0 \ldots 360^0$	$360^0 \ldots 540^0$	$540^0 \ldots 720^0$
chord A	$0 \ldots 2$	$2 \ldots 0$	$0 \ldots -2$	$-2 \ldots 0$

We have already seen that $\sin A = \sin (180^0 - A)$ and that $\cos A = -\cos (180^0 - A)$ (Art. 14); we shall now prove more general propositions of the same kind.

18. *To prove that,* n *being any integer,*

$$\sin A = \pm \sin (4n . 90^0 \pm A) = \pm \sin \{(4n + 2)\, 90^0 \mp A\}.$$

It is manifest that none of the trigonometrical functions, (except the chord,) are altered by supposing the angle to be increased by any number of complete revolutions of the revolving line; that is to say, $\sin A = \sin (4n . 90^0 + A)$.



Again, if we take $P'OA = -POA$, it is evident that $P'N = PN$, and

$$\therefore \sin A = -\sin(-A)$$
$$= -\sin(4n \cdot 90^0 - A)$$

by what precedes.

Again, if we produce PO, $P'O$ to P''', P'', it may be seen that

$$\sin(180^0 - A) = \sin A,$$

and $$\sin(180^0 + A) = -\sin A.$$

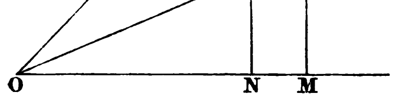
Hence $$\sin A = \pm \sin(2.90^0 \mp A)$$
$$= \pm \sin \{(4n+2) \cdot 90^0 \mp A\}.$$

Other propositions may be demonstrated in like manner: as for example,

$$\cos A = \cos(720^0 \pm A),$$
$$\cos A = -\cos(540^0 \pm A),$$
$$\tan A = \pm \tan(180^0 \pm A).$$

ON FORMULÆ INVOLVING MORE THAN ONE ANGLE.

19. *Given the sines and cosines of two angles, to find the sine and cosine of their sum or difference.*



Let POQ, QOM be any two angles, which call A and B respectively.

From any point P in OP draw PQ perpendicular to OQ, and from P, Q draw PN, QM perpendicular to OM, and QR perpendicular to PN.

It will be seen that

$$QPR = 90^0 - PQR = RQO = QOM = B.$$

Then $\sin (A + B) = \sin PON = \dfrac{PN}{OP} = \dfrac{PR + QM}{OP}$

$$= \frac{PQ}{OP} \cdot \frac{PR}{PQ} + \frac{OQ}{OP} \cdot \frac{QM}{OQ}$$

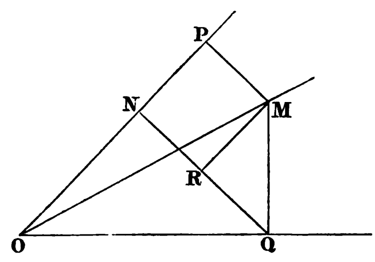
$$= \sin A \cos B + \cos A \sin B \;\ldots\ldots\; (1).$$

Again, $\cos (A + B) = \cos PON = \dfrac{ON}{OP} = \dfrac{OM - QR}{OP}$

$$= \frac{OQ}{OP} \cdot \frac{OM}{OQ} - \frac{PQ}{OP} \cdot \frac{QR}{PQ}$$

$$= \cos A \cos B - \sin A \sin B \;\ldots\ldots\; (2).$$

For the sine and cosine of the *difference* of two angles we must make a new construction.



Let $POQ = A$, $QOM = B$ as before, then $POM = A - B$. From any point M in OM draw the lines MP, MQ perpendicular to OP and OQ respectively, QN perpendicular to OP, and MR to QN. It is evident that $MQR = A$.

Then $\sin(A-B) = \sin POM = \dfrac{PM}{OM} = \dfrac{QN - QR}{OM}$

$$= \frac{QN}{OQ} \cdot \frac{OQ}{OM} - \frac{QR}{MQ} \cdot \frac{MQ}{OM}$$

$$= \sin A \cos B - \cos A \sin B \;\ldots\ldots (3).$$

Again, $\cos(A-B) = \cos POM = \dfrac{OP}{OM} = \dfrac{ON + MR}{OM}$

$$= \frac{ON}{OQ} \cdot \frac{OQ}{OM} + \frac{MR}{MQ} \cdot \frac{MQ}{OM}$$

$$= \cos A \cos B + \sin A \sin B \;\ldots\ldots (4).$$

20. It will be observed, that the preceding formulæ have been proved by means of figures which suppose both A and B, as well as $A+B$ and $A-B$, to be each less than a right angle; nevertheless we are justified in concluding that the same formulæ will hold in all cases, provided the proper signs be given to the quantities which enter: and herein consists one great advantage of the mode of denoting the directions of lines by their signs, that when any formula has been established for a standard case in which all the lines are positive, the same may be safely assumed to be true in all other cases.

As an example of what has been here remarked, it may be observed, that the formulæ (3) and (4) just proved, may be deduced from the formulæ (1) and (2) by changing the sign of the angle B. For we have from (1),

$$\sin(A+B) = \sin A \cos B + \cos A \sin B;$$

now write $-B$ for B, and we have

$$\sin(A-B) = \sin A \cos(-B) + \cos A \sin(-B)$$

$$= \sin A \cos B - \cos A \sin B,$$

which agrees with (3); and similarly (4) may be deduced from (2).

But still further, (2) may be deduced from (1): for we have

$$\cos(A + B) = \sin\{(90^0 - A) - B\}$$
$$= \sin(90^0 - A)\cos(-B) + \cos(90^0 - A)\sin(-B)$$
$$= \cos A \cos B - \sin A \sin B.$$

Hence it appears, that the only formula which it is quite necessary to establish by reference to a geometrical figure, is the fundamental formula (1).

21. By making $B = A$ we obtain the following formulæ:

$$\sin 2A = 2 \sin A \cos A,$$
$$\cos 2A = \cos^2 A - \sin^2 A,$$

which last formula, in consequence of the relation

$$\cos^2 A + \sin^2 A = 1,$$

may be put under either of the following forms,

$$\cos 2A = 2\cos^2 A - 1,$$
$$\cos 2A = 1 - 2\sin^2 A.$$

Also, conversely, we have the following useful formulæ,

$$\cos A = \sqrt{\frac{1 + \cos 2A}{2}},$$

$$\sin A = \sqrt{\frac{1 - \cos 2A}{2}}.$$

22. *To express tan* (A ± B) *in terms of tan* A *and tan* B.

We have

$$\tan(A \pm B) = \frac{\sin(A \pm B)}{\cos(A \pm B)}$$

$$= \frac{\sin A \cos B \pm \cos A \sin B}{\cos A \cos B \mp \sin A \sin B}$$

(dividing numerator and denominator by $\cos A \cos B$)

$$= \frac{\tan A \pm \tan B}{1 \mp \tan A \tan B}.$$

Making $B = A$ we obtain the formula,

$$\tan 2A = \frac{2 \tan A}{1 - \tan^2 A}.$$

23. The following formulæ are of great service in trigonometrical investigations.

We have

$$A = \frac{A+B}{2} + \frac{A-B}{2},$$

$$B = \frac{A+B}{2} - \frac{A-B}{2};$$

$$\therefore \sin A = \sin \frac{A+B}{2} \cos \frac{A-B}{2} + \cos \frac{A+B}{2} \sin \frac{A-B}{2},$$

$$\sin B = \sin \frac{A+B}{2} \cos \frac{A-B}{2} - \cos \frac{A+B}{2} \sin \frac{A-B}{2},$$

$$\cos A = \cos \frac{A+B}{2} \cos \frac{A-B}{2} - \sin \frac{A+B}{2} \sin \frac{A-B}{2},$$

$$\cos B = \cos \frac{A+B}{2} \cos \frac{A-B}{2} + \sin \frac{A+B}{2} \sin \frac{A-B}{2}.$$

Hence

$$\sin A + \sin B = 2 \sin \frac{A+B}{2} \cos \frac{A-B}{2} \quad \ldots\ldots \ (1)$$

$$\sin A - \sin B = 2 \cos \frac{A+B}{2} \sin \frac{A-B}{2} \quad \ldots\ldots \ (2)$$

$$\cos A + \cos B = 2 \cos \frac{A+B}{2} \cos \frac{A-B}{2} \quad \ldots\ldots \ (3)$$

$$\cos B - \cos A = 2 \sin \frac{A+B}{2} \sin \frac{A-B}{2} \quad \ldots\ldots \ (4)$$

24. The formulæ which have been investigated in the last few articles may be easily extended to functions of *three* or more angles. For instance,

$$\sin(A+B+C) = \sin(A+B)\cos C + \cos(A+B)\sin C$$

$$= (\sin A\cos B + \cos A\sin B)\cos C + (\cos A\cos B - \sin A\sin B)\sin C$$

$$= \sin A\cos B\cos C + \sin B\cos A\cos C$$
$$+ \sin C\cos A\cos B - \sin A\sin B\sin C.$$

Similarly we may express $\cos(A+B+C)$, $\tan(A+B+C)$, &c., in terms of the trigonometrical functions of the simple angles.

In like manner we may express $\sin 3A$, $\cos 3A$.

$$\sin 3A = \sin(A+2A) = \sin A\cos 2A + \cos A\sin 2A$$
$$= \sin A(1 - 2\sin^2 A) + 2\sin A\cos^2 A$$
$$= \sin A\left\{1 - 2\sin^2 A + 2 - 2\sin^2 A\right\}$$
$$= 3\sin A - 4\sin^3 A;$$

$$\cos 3A = \cos(A+2A) = \cos A\cos 2A - \sin A\sin 2A$$
$$= \cos A(2\cos^2 A - 1) - 2\sin^2 A\cos A$$
$$= \cos A\left\{2\cos^2 A - 1 - 2 + 2\cos^2 A\right\}$$
$$= 4\cos^3 A - 3\cos A.$$

DETERMINATION OF THE NUMERICAL VALUES OF THE TRIGONOMETRICAL FUNCTIONS OF ANGLES.

25. The value of the trigonometrical functions of certain angles may be expressed with great facility and simplicity, that of others must be found by a laborious course of calculation. A few which present the most simple investigation are here given.

To find the value of sin 45°.

In general $\sin^2 A + \cos^2 A = 1$.

$$\text{Let } A = 45^\circ; \quad \therefore \cos A = \cos 45^\circ = \sin 45^\circ;$$

$$\therefore 2\sin^2 45^\circ = 1,$$

$$\sin^2 45^\circ = \tfrac{1}{2},$$

$$\sin 45^\circ = \frac{1}{\sqrt{2}};$$

the positive sign of the radical is taken, because we know that sin 45° is positive.

Hence also $\cos 45^\circ = \dfrac{1}{\sqrt{2}}$, and $\tan 45^\circ = \cot 45^\circ = 1$.

To find the value of sin 30°.

$$\cos 30^\circ = \sin 60^\circ = 2\sin 30^\circ \cos 30^\circ;$$

$$\therefore 2\sin 30^\circ = 1,$$

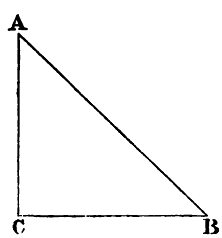
$$\sin 30^\circ = \tfrac{1}{2}.$$

Also $\cos 30^\circ = \sqrt{1 - \sin^2 30^\circ} = \dfrac{\sqrt{3}}{2}$,

$$\tan 30^\circ = \frac{1}{\sqrt{3}}.$$

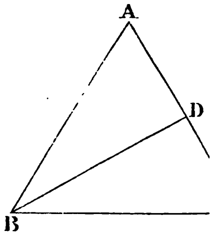
The values of sin 45° and sin 30° may be obtained very simply by a geometrical figure, thus:

Let ACB be an isosceles triangle and having a right angle at C. Then each of the angles A and B is an angle of 45°.



$$\therefore \sin 45^0 = \frac{BC}{AB} = \frac{BC}{\sqrt{AC^2 + BC^2}} = \frac{BC}{\sqrt{2BC^2}} = \frac{1}{\sqrt{2}}.$$

Again, let ABC be an equilateral triangle, and draw BD perpendicular to, and therefore bisecting AC. Then since the angles A, B, C, are equal to each other and together equal to 180^0, each of them is 60^0, and therefore ABD, which manifestly is half of ABC, $= 30^0$.



$$\therefore \sin 30^0 = \frac{AD}{AB} = \frac{1}{2}.$$

To find the value of sin 18^0.

$$36^0 = 90^0 - 54^0;$$

$$\therefore \sin 36^0 = \cos 54^0.$$

We have seen (Art. 24) that $\cos 3A = 4\cos^3 A - 3\cos$ also, $36^0 = 2 \times 18^0$, and $54^0 = 3 \times 18^0$;

$$\therefore 2 \sin 18^0 \cos 18^0 = 4 \cos^3 18 - 3 \cos 18^0,$$

$$2 \sin 18^0 = 4 \cos^2 18^0 - 3 = 1 - 4 \sin^2 18^0,$$

$$\sin^2 18^0 + \frac{\sin 18^0}{2} = \frac{1}{4}.$$

Completing the square,

$$\sin^2 18^0 + \frac{\sin 18^0}{2} + \frac{1}{16} = \frac{1}{4} + \frac{1}{16} = \frac{5}{16},$$

$$\sin 18^0 = \frac{-1 + \sqrt{5}}{4};$$

the positive sign is given to the radical, because we know t sin 18^0 is positive.

By means of the preceding values, we may investigate a variety of others. For instance, if it were required to find sin 12^0, we should have

$$\sin 12^0 = \sin (30^0 - 18^0),$$

$$= \sin 30^0 \cos 18^0 - \cos 30^0 \sin 18^0;$$

and consequently sin 12^0 becomes known. In short, we can find the values of the trigonometrical functions of any angles, which are combinations of those which have been discussed.

26. The values of the trigonometrical functions of all angles from $1'$ up to 90^0 may be calculated by methods not here explained, and arranged in tables for the purposes of practical application. Still more useful are tables containing the *logarithms* of these values; the student who is desirous of understanding the mode of constructing such tables is referred to the treatise of *Snowball* or *Hymers*, or any other complete treatise on Trigonometry.

ON THE SOLUTION OF TRIANGLES.

27. A triangle consists of six parts, viz., three sides and three angles: three of these being given, the others may be found, unless the three *angles* be the three given parts, in which case nothing further can be found. The angles of a triangle are not in fact *independent parts*, their sum being always two right angles; it may be said that if three *independent parts* be given, the others may be found.

28. Before we engage ourselves with the solution of a triangle we must investigate certain formulæ connecting the parts of a triangle.

We shall denote the angles of a triangle by A, B, C, the sides respectively opposite to them by a, b, c.

29. *The sides of a triangle are proportional to the sines of the opposite angles.*

FIG. 1.

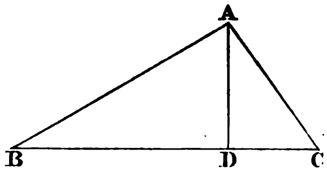
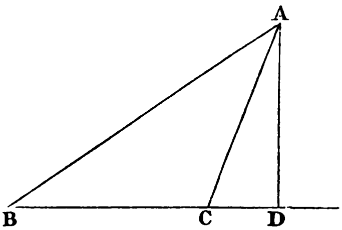


FIG. 2.



Let ABC be a triangle. From any angular point A let fall the perpendicular AD on the opposite side (fig. 1), or that side produced (fig. 2).

Then in the first case, $AD = c \sin B = b \sin C$;

in the second, $AD = c \sin B = b \sin (180^0 - C) = b \sin C$;

$$\text{therefore in both, } \frac{\sin B}{b} = \frac{\sin C}{c}.$$

In like manner,

$$\frac{\sin A}{a} = \frac{\sin B}{b} = \frac{\sin C}{c};$$

which proves the theorem.

30. *To express the cosine of an angle of a triangle in terms of the sides.*

With the same figures as in the last proposition, we have in fig. 1,

$$a = BD + DC = c \cos B + b \cos C,$$

in fig. 2,

$$a = BD - DC = c \cos B - b \cos (180^0 - C) = c \cos B + b \cos C;$$

therefore in both cases,

$$\frac{\cos B}{b} + \frac{\cos C}{c} = \frac{a^2}{abc},$$

in like manner, $$\frac{\cos C}{c} + \frac{\cos A}{a} = \frac{b^2}{abc},$$

and $$\frac{\cos A}{a} + \frac{\cos B}{b} = \frac{c^2}{abc};$$

adding the last two of these equations and subtracting the first, we have

$$\cos A = \frac{b^2 + c^2 - a^2}{2bc},$$

which is the expression required.

31. *To express* $\sin\frac{A}{2}$, $\cos\frac{A}{2}$, $\tan\frac{A}{2}$ *in terms of the sides.*

By Art. 21, $$\sin^2\frac{A}{2} = \frac{1-\cos A}{2} = \frac{2bc - b^2 - c^2 + a^2}{4bc},$$

$$= \frac{a^2-(b-c)^2}{4bc} = \frac{(a-b+c)(a+b-c)}{4bc}.$$

Let $a + b + c = 2S$,

$$\therefore\ a - b + c = 2(S - b),$$

$$a + b - c = 2(S - c),$$

and $$\sin\frac{A}{2} = \sqrt{\frac{(S-b)(S-c)}{bc}}.$$

In like manner,

$$\cos^2\frac{A}{2} = \frac{1 + \cos A}{2} = \frac{2bc + b^2 + c^2 - a^2}{4bc},$$

$$= \frac{(b+c)^2 - a^2}{4bc} = \frac{(a+b+c)(b+c-a)}{4bc},$$

$$= \frac{S(S-a)}{bc};$$

$$\therefore \cos\frac{A}{2} = \sqrt{\frac{S(S-a)}{bc}}.$$

Hence also we have

$$\tan\frac{A}{2} = \sqrt{\frac{(S-b)(S-c)}{S(S-a)}}.$$

32. *To express* sin A *in terms of the sides.*

This is done at once by means of the expressions just proved: for

$$\sin A = 2\sin\frac{A}{2}\cos\frac{A}{2},$$

$$= \frac{2}{bc}\sqrt{S(S-a)(S-b)(S-c)}.$$

This expression has the advantage of being adapted to *logarithmic computation*, that is to say, it consists of *factors*. The formula for the cosine (Art. 30) has not this advantage.

33. *To prove the following formula,*

$$\tan\frac{\text{A}-\text{B}}{2} = \frac{\text{a}-\text{b}}{\text{a}+\text{b}}\cot\frac{\text{C}}{2}.$$

We have by Art. 29,

$$\frac{a}{b} = \frac{\sin A}{\sin B};$$

$$\therefore \frac{a-b}{a+b} = \frac{\sin A - \sin B}{\sin A + \sin B}$$

$$= \frac{2\cos\frac{A+B}{2}\sin\frac{A-B}{2}}{2\sin\frac{A+B}{2}\cos\frac{A-B}{2}}, \text{ by Art. 23,}$$

$$= \frac{\tan\frac{A-B}{2}}{\tan\frac{A+B}{2}};$$

but $A + B = 180^0 - C$,

$$\therefore \frac{A+B}{2} = 90^0 - \frac{C}{2},$$

$$\text{and } \tan\frac{A+B}{2} = \cot\frac{C}{2}.$$

Substituting in the equation before obtained, we have

$$\tan\frac{A-B}{2} = \frac{a-b}{a+b}\cot\frac{C}{2}.$$

Having established the preceding necessary formulæ, we now proceed to the solution of triangles.

34. *Let two angles and the side between them* (A, C, b) *be given*

The other angle is known at once, because

$$B = 180^0 - A - C.$$

Again, we have

$$a = b\frac{\sin A}{\sin B},$$

$$\text{and } c = b\frac{\sin C}{\sin B};$$

which determine a and c.

35. *Let two angles and a side opposite to one of them* (A, C, a) *be given.*

As before, the third angle is known, because

$$B = 180^0 - A - C.$$

Also,
$$b = a\frac{\sin B}{\sin A},$$
$$c = a\frac{\sin C}{\sin A},$$
which determine b and c.

36. *Let two sides and the included angle* (C, a, b) *be given.*

We determine the other angles thus,
$$A + B = 180^0 - C.$$

Again, by Art. 33,
$$\tan\frac{A-B}{2} = \frac{a-b}{a+b}\cot\frac{C}{2},$$

which determines $A - B$: thus $A + B$ and $A - B$ are both known; and therefore A and B, which are $\frac{A+B}{2} + \frac{A-B}{2}$ and $\frac{A+B}{2} - \frac{A-B}{2}$ respectively, are also known.

To determine c we have
$$c = a\frac{\sin C}{\sin A}.$$

37. There is another mode of solving the triangle in this case.

Since
$$\cos C = \frac{a^2 + b^2 - c^2}{2ab},$$
we have $c^2 = a^2 + b^2 - 2ab\cos C$;

and this equation in fact determines c, but in its present state it would be practically of no use because it is *not adapted to logarithmic computation*; we can however modify it in such a manner as to render it suitable, as follows:

$$c^2 = a^2 + b^2 - 2ab\cos C$$

$$= (a^2 + b^2)\left(\cos^2\frac{C}{2} + \sin^2\frac{C}{2}\right) - 2ab\left(\cos^2\frac{C}{2} - \sin^2\frac{C}{2}\right)$$

$$= (a - b)^2\cos^2\frac{C}{2} + (a + b)^2\sin^2\frac{C}{2}$$

$$= (a - b)^2\cos^2\frac{C}{2}\left\{1 + \left(\frac{a + b}{a - b}\right)^2\tan^2\frac{C}{2}\right\}.$$

Since the tangent of an angle may be of any magnitude, there will be an angle the tangent of which is $\frac{a + b}{a - b}\tan\frac{C}{2}$; let θ be such an angle, so that

$$\tan\theta = \frac{a + b}{a - b}\tan\frac{C}{2},$$

or that $\log\tan\theta = \log(a + b) - \log(a - b) + \log\tan\frac{C}{2}$.

The value of θ is found by looking in the tables and finding an angle which has for its logarithmic tangent the preceding quantity, and may therefore be now supposed to be *known*.

Also we have

$$c^2 = (a - b)^2\cos^2\frac{C}{2}(1 + \tan^2\theta)$$

$$= (a - b)^2\cos^2\frac{C}{2}\sec^2\theta,$$

$$\text{or}\quad c = (a - b)\frac{\cos\frac{C}{2}}{\cos\theta},$$

which equation determines c.

The sides a, b, c being all known, A and B may be determined by any of the expressions given in Arts. (30), (31), (32).

38. The process of adapting a formula to logarithmic computation, which has been introduced in the preceding article, is one of frequent use. The angle θ, which has been employed to assist the calculation, is called a *subsidiary angle.* It will be seen that the possibility of facilitating calculations by this means, arises from the fact of our possessing tables containing all the trigonometrical functions of the same angle, so that as soon as one function of an angle is known all the others become known; for instance, in the case we have been considering, as soon as a certain quantity was fixed upon as representing the *tangent* of an angle, the *cosine* of that angle was known by inspection of the table, and the calculation was spared by which it would have been necessary to determine the cosine from the tangent. The adaptation of formulæ to logarithmic computation is a matter not of rule, but of ingenuity, and frequently a formula may be adapted in various ways equally good; we must however be careful to ascertain that the supposition we make involves no absurdity; for instance, we may not assume a quantity to be $= \cos\theta$, (θ being the subsidiary angle,) unless we have ascertained that the quantity is not greater than unity, and so in other instances.

39. We may here make a remark respecting the tables of logarithmic trigonometrical functions, which, if we had troubled the student with a complete account of the formation of such tables, would have found a more fitting place elsewhere.

It appeared from the account of logarithms given in the treatise on Algebra, (page 80, Art. 126,) that the logarithms of numbers less than unity have *negative characteristics.* Now all sines and cosines are less than unity, (except $\sin 90^0$ and $\cos 0^0$,) and therefore their logarithms have negative characteristics; but it is not convenient to register such quantities in tables, and therefore it is usual to add 10 to each of the logarithmic functions, and thus the characteristic -1 is replaced

by 9. This is mere matter of convenient arrangement, but it entails the following precaution, that when the logarithms of numbers and those of trigonometrical functions occur in the same equation, 10 *must be subtracted from each logarithmic function of an angle.*

For instance, suppose we had the equation

$$b = a \sin C,$$

in which a and C are given and b is to be found; we should take the logarithms thus,

$$\log b = \log a + \log \sin C - 10,$$

instead of

$$\log b = \log a + \log \sin C.$$

Or we may make the distinction between the two kinds of logarithms more manifest by a difference of notation, and may write the preceding formula thus,

$$\log_{10} b = \log_{10} a + L \sin C - 10,$$

where $\log_{10}$ indicates the logarithm of a number to base 10, and L the logarithm of a trigonometrical function to the same base when 10 has been added to it.

40. *Let two sides and an angle opposite to one of them* (a, b, A) *be given.*

To determine B, we have

$$\sin B = \frac{b}{a} \sin A;$$

C is then known from the formula

$$C = 180^0 - A - B,$$

and c from

$$c = a \frac{\sin C}{\sin A}.$$

41. The solution of the triangle in this case however is not without ambiguity; for the equation

$$\sin B = \frac{b}{a} \sin A,$$

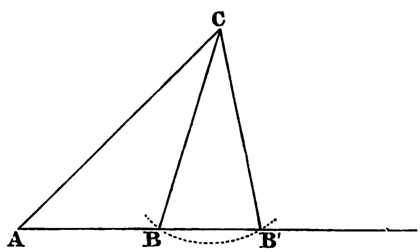
does not determine *one* angle but *two*, because

$$\sin B = \sin (180^0 - B),$$

and the question is whether there is any test to guide us in choosing one of the values rather than the other.

Now $180^0 - B = A + C$, and therefore B and $180^0 - B$ cannot both be less than A; but the greater side is opposite the greater angle, (Euclid, I. 18,) consequently, if b be *less* than a, B must be less than A, but the two values of B determined cannot both be less than A, therefore we know which to choose. On the other hand, if b be *greater* than a, B must be greater than A, but both of the values determined may be so, therefore the solution is *ambiguous*.

42. The same results may be obtained very simply by geometrical considerations.



Let CAB be the given angle, AC the given side; with centre C and distance a (the value of the other given side) describe an arc of a circle, which, if a be less than b will (as in the figure) cut the straight line AB in two points B, B' *on the same side of* A. Now each of the triangles CAB, CAB', has all the data of the question, and therefore the solution is ambiguous. If a had been greater than b, the points BB' would have been on opposite sides of A, and there would have been only one triangle answering the given conditions.

43. It may perhaps be not without use to give a third investigation of *the ambiguous case* in the solution of triangles.

We have the formula,

$$a^2 = b^2 + c^2 - 2bc \cos A,$$

$$\text{or } c^2 - 2bc \cos A = a^2 - b^2.$$

In this equation ab and A are given, and we may therefore find c; completing the square,

$$c^2 - 2bc \cos A + b^2 \cos^2 A = a^2 - b^2 \sin^2 A;$$

$$\therefore\ c = b \cos A \pm \sqrt{a^2 - b^2 \sin^2 A}.$$

We have here *two* values of c, and if both values are admissible the solution is ambiguous; but the only thing which can limit their admissibility is their *sign*; hence the solution is ambiguous if both values of c are *positive*,

$$i.\,e. \text{ if } b \cos A > \sqrt{a^2 - b^2 \sin^2 A},$$

$$\text{or } b^2 \cos^2 A > a^2 - b^2 \sin^2 A,$$

$$\text{or } b^2 (\cos^2 A + \sin^2 A) > a^2;$$

$$\text{or } \quad b > a;$$

which is the same conclusion as that which we have arrived at before.

44. *Let all the sides be given.*

This case may be solved by any of the formulæ,

$$\sin \frac{A}{2} = \sqrt{\frac{(S - b)(S - c)}{bc}},$$

$$\cos \frac{A}{2} = \sqrt{\frac{S(S - a)}{bc}},$$

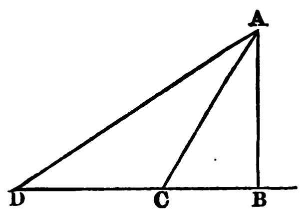
$$\tan\frac{A}{2} = \sqrt{\frac{(S-b)(S-c)}{S(S-a)}},$$

$$\sin A = \frac{2}{bc}\sqrt{S(S-a)(S-b)(S-c)}$$

45. We have now considered all the cases of oblique-angled triangles. In practice, the triangles to be solved are frequently right-angled, in which case the solution is much simplified. One application of the methods of solving triangles is to the finding of the heights and distances of inaccessible objects; in problems of this kind we suppose the magnitudes of certain lines and angles to be measured by means of proper instruments, a description of which however would not be appropriate here. We subjoin a few simple examples of the method of

FINDING HEIGHTS AND DISTANCES.

Ex. 1. A river of unknown breadth runs between an observer and a tower on its opposite bank; find the breadth of the river and the height of the tower.



Let AB be the tower, BC the breadth of the river; let the angle BCA be observed, and then let the observer retreat in the direction of the line BC to D, measuring the distance CD, and observing the angle CDA.

$$\text{Let } ACB = \alpha,\quad CDA = \beta,\quad CD = a,$$

$$AB = x,\qquad BC = y.$$

Then $y = x\cot\alpha$ from triangle ABC,

and $y + a = x\cot\beta$ ....ABD,

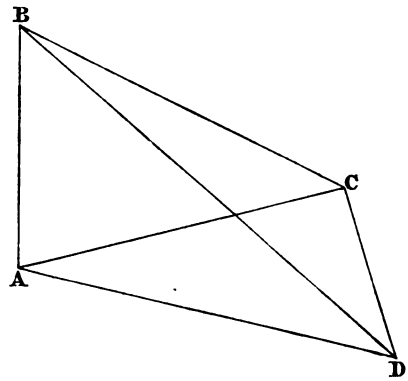
subtracting the first of these equations from the second

$$a = x(\cot\beta - \cot\alpha);$$

$$\therefore x = \frac{a}{\cot\beta - \cot\alpha} = a\frac{\sin\alpha\sin\beta}{\sin(\alpha-\beta)},$$

$$y = x\cot\alpha = a\frac{\cos\alpha\sin\beta}{\sin(\alpha-\beta)}.$$

Ex. 2. From the top of a tower, a person observes the angle of depression of two distant points in the horizontal plane, the distance of which from each other he knows, and also the angle subtended at his eye by the line joining the two points; required the height of the tower.



Let AB be the tower, C, D the two points; then the angles observed are BCA, BDA, CBD, which call α, β, γ respectively; also let $CD = a$, and let the height of the tower $= x$.

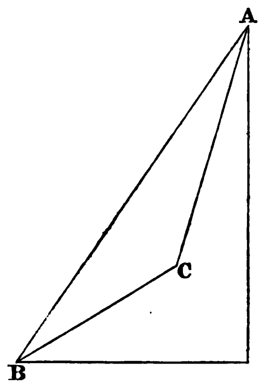
Then in the triangle BCD, $BC = x \operatorname{cosec}\alpha$, $BD = x\operatorname{cosec}\beta$;

$$\therefore a^2 = x^2\operatorname{cosec}^2\alpha + x^2\operatorname{cosec}^2\beta - 2x^2\operatorname{cosec}\alpha\operatorname{cosec}\beta\cos\gamma,$$

and $$x = \frac{a}{\sqrt{\operatorname{cosec}^2\alpha + \operatorname{cosec}^2\beta - 2\operatorname{cosec}\alpha\operatorname{cosec}\beta\cos\gamma}}.$$

Ex. 3. From a station B at the base of a mountain, its summit A is seen at an elevation of 60°; after walking one mile towards the summit up a plane making an angle of 30°

with the horizon, to another station C, the angle BCA is ob served to be 135^0. Find the height of the mountain.



From the triangle ABC, we have

$$AB = BC\frac{\sin BCA}{\sin BAC},$$

$$= BC \cdot \frac{\sin BCA}{\sin (BCA + ABC)} = BC \cdot \frac{\sin 135^0}{\sin (135^0 + 30^0)}$$

$$= BC \cdot \frac{\sin 45^0}{\sin 15^0},$$

and the height of the mountain $= AB \sin 60^0$

$$= BC \cdot \frac{\sin 60^0 \sin 45^0}{\sin 15^0}.$$

But $\sin 60^0 = \frac{\sqrt{3}}{2}$, $\quad \sin 45^0 = \frac{1}{\sqrt{2}}$,

$$\sin 15^0 = \sqrt{\frac{1 - \cos 30^0}{2}} = \sqrt{\frac{2 - \sqrt{3}}{4}} = \frac{1}{2}\sqrt{\frac{4 - 2\sqrt{3}}{2}} = \frac{\sqrt{3} - 1}{2\sqrt{2}}$$

and $BC = 1760$ yards;

$$\therefore \text{ the height} = 1760 \cdot \frac{\sqrt{3}}{\sqrt{3} - 1} = 880 (3 + \sqrt{3}),$$

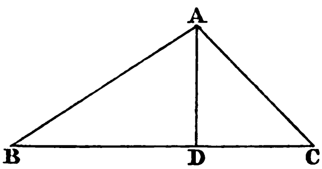
$= 880 \times 4.7$ nearly

$= 2136$ yards.

We shall here subjoin a few propositions relating to triangles.

46. *To find the area of a triangle in terms of the sides.*

Let ABC be the triangle; from A draw the perpendicular AD on the opposite side BC. Then

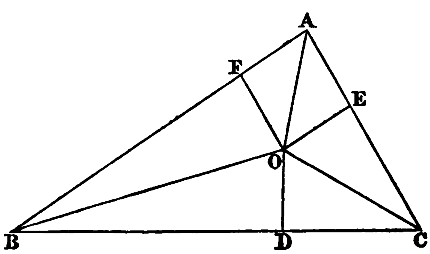


$$\text{the area} = \frac{BC \times AD}{2} = \frac{a \times b \sin C}{2};$$

but (by Art. 32) $\sin C = \dfrac{2}{ab}\sqrt{S(S-a)(S-b)(S-c)}$;

therefore the area $= \sqrt{S(S-a)(S-b)(S-c)}$.

47. *To find the radius of a circle inscribed in a triangle.*



Let ABC be the triangle; bisect the angles, and let O be the point in which the bisecting lines meet, then (by Euclid IV. 4), O is the centre, and if we draw OD, OE, OF, perpendicular to the sides BC, AC, AB, respectively, any one of these

will be the radius of the inscribed circle. Call the radius r; then

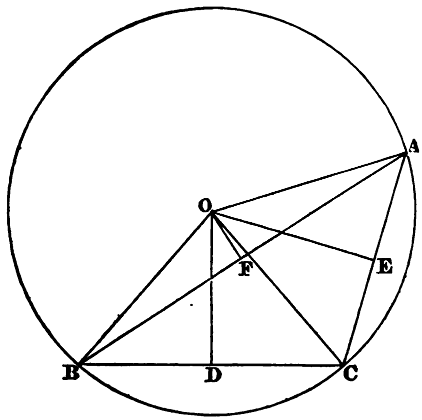
$$\text{area of } \triangle ABC = \triangle BOC + \triangle AOC + \triangle AOB,$$

$$\text{or } \sqrt{S(S-a)(S-b)(S-c)} = \frac{ra}{2} + \frac{rb}{2} + \frac{rc}{2} = rS;$$

$$\therefore r = \sqrt{\frac{(S-a)(S-b)(S-c)}{S}}.$$

48. *To find the radius of a circle circumscribed about a triangle.*

Bisect the sides in the points D, E, F, and from the points of bisection draw perpendiculars meeting in the point



O; then (by Euclid IV. 5), O is the centre of the circumscribed circle; join OA, OB, OC, and describe the circle BCA. Call the radius R; then

$$\text{(as in Art. 46) area of } ABC = \frac{AB \cdot BC}{2} \sin ABC,$$

$$\text{or } \sqrt{S(S-a)(S-b)(S-c)} = \frac{ac}{2} \sin \frac{AOC}{2}, \text{ (Euc. III. 20)},$$

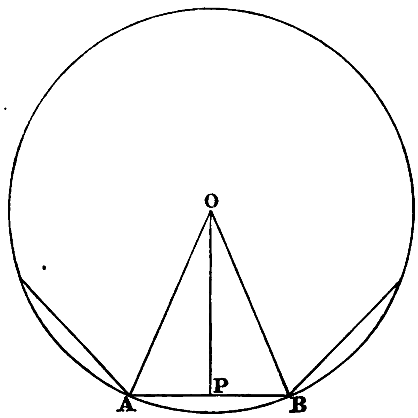
$$= \frac{ac}{2} \sin AOE$$

$$= \frac{ac}{2} \cdot \frac{b}{2R};$$

$$\therefore R = \frac{abc}{4\sqrt{S(S-a)(S-b)(S-c)}}.$$

49. *To find the circumference and area of a regular polygon inscribed in a circle.*

Let O be the centre of the circle, AB one of the sides of the polygon, of which suppose that the number is n;



and let the radius of the circle be r. Join OA, OB and draw OP perpendicular to AB; then the angle $AOB = \frac{360^0}{n}$, and

circumference of polygon $= n \,.\, AB$,

$$= 2n \,.\, AP = 2n\, AO \sin AOP,$$

$$= 2nr \sin \frac{180^0}{n}.$$

Again, area of polygon $= n \,.$ area of AOB,

$$= n\,\frac{AB \,.\, OP}{2},$$

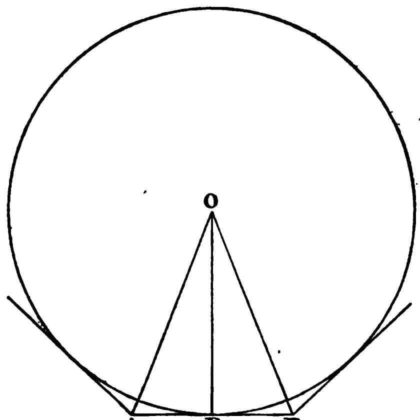
$$= n\, AO \sin AOP \times AO \cos AOP,$$

$$= nr^2 \sin \frac{180^0}{n} \cos \frac{180^0}{n}.$$

From these expressions we see, that if the number of sides is given, the *circumference* of the polygon is proportional to the *radius*, and the *area* to the *square of the radius* of the circumscribing circle.

50. *To find the circumference and area of a regular polygon circumscribed about a circle.*

Let AB be any one of the sides of the polygon, touching



the circle at P, n the number of sides, and r the radius, as before. Join OP. Then circumference of polygon

$$\begin{aligned} &= n \,.\, AB, \\ &= 2n\, AP, \\ &= 2n\, OP \tan AOP, \\ &= 2nr \tan\frac{180^0}{n}. \end{aligned}$$

Again, area of polygon $= n$. area of AOB,

$$\begin{aligned} &= n\frac{AB\,.\,OP}{2}, \\ &= nOP\tan AOP \times OP, \\ &= nr^2\tan\frac{180^0}{n}. \end{aligned}$$

As in the case of the inscribed polygon, we observe that the *circumference* is proportional to the *radius*, and the *area* to the *square of the radius*, the number of sides being given.

51. If we suppose a regular polygon to be inscribed in a circle, and another of the same number of sides to be circumscribed about it, it is easy to see, that the greater the number of sides, the more nearly will each of the polygons approximate to the other and to the circle which lies between them. In fact, suppose C, C' to be the circumferences of the two polygons, then we have seen that

$$C = 2nr \sin \frac{180^0}{n},$$

$$C' = 2nr \tan \frac{180^0}{n};$$

$$\therefore \frac{C}{C'} = \cos \frac{180^0}{n}.$$

Suppose n to be indefinitely great, then $\cos \frac{180^0}{n}$ becomes $\cos 0^0$ or 1, and $C = C'$.

Or let A, A' be the areas of the two polygons, then

$$A = nr^2 \sin \frac{180^0}{n} \cos \frac{180^0}{n},$$

$$A' = nr^2 \tan \frac{180^0}{n};$$

$$\therefore \frac{A}{A'} = \cos^2 \frac{180^0}{n},$$

and if n be indefinitely great, $A = A'$.

Consequently, if we suppose the number of the sides to be indefinitely increased, the inscribed and circumscribed polygon will coincide with each other, and therefore with the circle. We may therefore consider a circle as being a regular polygon having an indefinite number of sides, and may extend to it those properties which we have proved concerning polygons. Now we have seen that the circumference of polygons is pro-

portional to the radius, and the area to the square of the radius of the circle, whether inscribed or circumscribed, and therefore we conclude that the *circumference* of a circle is proportional to its *radius*, and the area to the *square of the radius*.

We have in fact, (taking the inscribed polygon,)

$$C = 2nr \sin\frac{180^0}{n}, \text{ and } A = nr^2 \sin\frac{180^0}{n}\cos\frac{180^0}{n};$$

if n be indefinitely great, $\cos\frac{180^0}{n} = 1$, and

$$C = 2\left(n\sin\frac{180^0}{n}\right)r, \quad A = \left(n\sin\frac{180^0}{n}\right)r^2.$$

What will be the value of $n\sin\frac{180^0}{n}$, when n is made indefinitely great? Its value, which is always denoted by π, might be found with some degree of accuracy by actually calculating the area of a polygon of a considerable number of sides; but this and other operose methods are superseded by modes of calculation of a more refined character, the introduction of which however would be unsuitable to the design of the present treatise: the result is that

$$\pi = 3.1415926535..............$$

The quantity π may be calculated to any degree of accuracy, but it is of the class of quantities called *incommensurable*, that is, it cannot be expressed by the ratio of any two whole numbers however great: in general we may consider 3.14159 as a sufficiently accurate value of π.

According to the notation we have adopted,

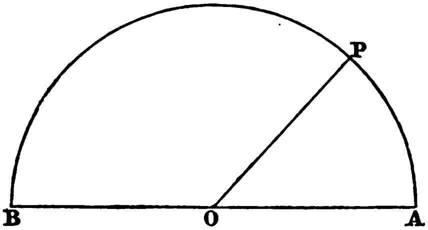
the circumference of a circle $= 2\pi r$,

... area .... $= \pi r^2$.

52. The introduction of the quantity π renders this a proper place for explaining another mode of measuring angles, besides that which has been hitherto used.

Hitherto we have considered the right angle to be divided into 90 degrees, and have measured angles by the number of degrees they contain; but there is another mode depending upon the proposition (Euc. VI. 33) that angles at the centre of a circle are proportional to the arcs on which they stand, and which is of frequent use.

Let POA be an angle at the centre O of a circle, the radius of which is r; APB a semicircle $= \pi r$; also let the length of the arc $AP = a$. Then, by Euclid,



$$\frac{\text{angle } POA}{2 \text{ right angles}} = \frac{a}{\pi r};$$

$$\therefore \text{ angle } POA = \frac{2 \text{ right angles}}{\pi} \cdot \frac{a}{r} \quad \text{(A)}.$$

Now supposing a and r to be given, although the angle POA will be determined, yet its *numerical value* will not be settled, unless we make some convention as to what angle we shall call unity. We are at liberty to make any convention that we please, but we shall be guided in our choice by the consideration of what will make the equation (A) the most simple, and it is manifest that the most simple form will be given to that equation by making

$$\frac{2 \text{ right angles}}{\pi} = 1. \quad \text{(B)},$$

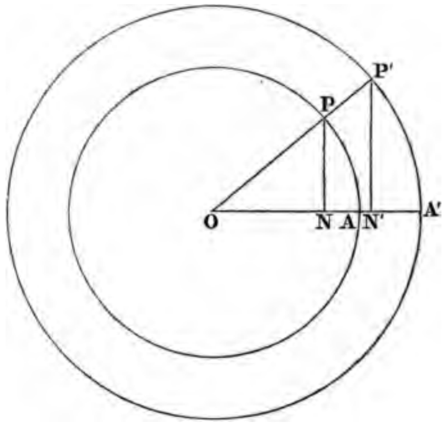
we shall then have, (denoting the numerical value of the angle POA by θ),

$$\theta = \frac{a}{r} \quad \text{(C)}.$$

Let us consider the results of the assumption (B). The numerical value of two right angles is the quantity π, instead of 180^0, as in the former method, and the unit of angle instead of being the ninetieth part of a right angle is $\dfrac{2 \text{ right angles}}{\pi}$ or $57^0\ 17'\ 44''\ 48'''$ nearly.

Again, making $\theta = 1$ in equation (C), we have $a = r$, which shews that *the unit of angle is that angle which is subtended by an arc of length equal to radius.*

53. Another mode of considering this subject is the following. Let POA be any angle, and about O as centre suppose any two circles described; let PA, $P'A'$ be the subtending arcs in the two circles, and draw PN, $P'N'$ perpendi-



cular to OAA'; then, if the radius of the circle were given, the arc PA would be a proper measure of the angle, and we might define PA to be the arc, PN the sine, ON the cosine, &c. of the angle POA; but this being inconvenient, in consequence of its being necessary to know the radius, we have defined $\dfrac{PN}{PO}$ to be the sine, $\dfrac{ON}{PO}$ to be the cosine, &c. of POA, these ratios being independent of the radius, since $\dfrac{PN}{PO} = \dfrac{P'N'}{P'O}$ and $\dfrac{ON}{PO} = \dfrac{ON'}{P'O}$; and on the same principle we should take as

the measure of the angle, not the arc PA, but $\frac{PA}{PO}$. Thus we are led, in a rather different way from that pursued in the last article, to choose $\frac{\text{arc}}{\text{radius}}$ as the measure of an angle, and this choice implies that the numerical value of two right angles is π, which was our first assumption in the other case.

From the equation (C)

$$\theta = \frac{a}{r},$$

we see, that if $r = 1$, $\theta = a$, that is to say, if we suppose the radius of the circle to be unity, the numerical value of the angle and the subtending arc is the same. Hence, if we make this supposition respecting the radius, we are not under the necessity of making any distinction between *arcs* and *angles*, since their numerical value is the same.

54. It is frequently a matter of indifference which mode of measuring angles we adopt, but this must be carefully borne in mind, that in every example either the one or the other must be used exclusively. It will perhaps be found generally advantageous to use that last explained, or the *circular measure* as it is sometimes called, as being the more brief.

It is easy to pass from one mode of measurement to the other: for suppose that θ is the circular measure of an angle, then the angle contains $\frac{\theta}{\pi}180$ degrees; and, conversely, if an angle contains n^0, its circular measure is $\frac{n}{180}\pi$.

CONIC SECTIONS.

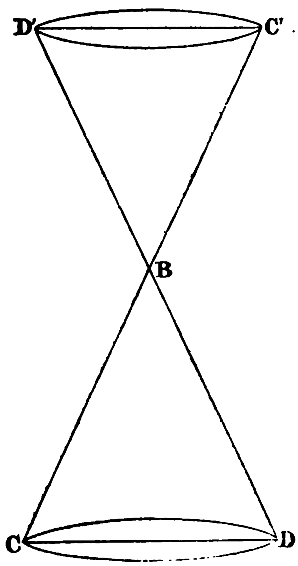
CONIC SECTIONS.

Definitions. A *right cone* is a surface generated by an indefinite straight line, which always passes through a given point, and makes a given angle with a given straight line passing through that point.

The point through which the generating line always passes, is called the *vertex* of the cone. The given straight line passing through the vertex, is called the *axis* of the cone.

The common notion of a cone is that of a pyramid standing on a circular base; it is clear that a cone as above defined will consist of two such pyramids of indefinite height, having their axes in the same straight line and their vertices coincident.

If we conceive a cone to be cut by a plane, the curve formed by the intersection will be different according to the position of the cutting plane. There are however only *three* different modes in which it is possible for the intersection to take place.



For distinctness of conception, let the annexed figure represent a cone; B is the vertex, $CDBC'D'$ is the intersection of the cone by the plane of the paper; the cone is supposed to be of indefinite length both above and below B. Then

(1) The cutting plane may be parallel to the line BC, in which case it will only cut one portion of the cone as BCD, and not the other $BC'D'$, and the curve formed by the intersection will evidently be a curve of one branch and unlimited in extent, since the cone is supposed to be unlimited. This curve is called the *parabola*.

(2) The cutting plane may be

inclined to the line BC, and may cut the cone wholly on one side of B, that is, may cut the portion BCD without cutting the portion $BC'D'$; in this case the curve will be one of limited extent, and of an oval form. This is the *ellipse*.

(3) The cutting plane may, as in the last case, be inclined to BC, but may cut the cone on both sides of B, that is, may cut the portion $BC'D'$ as well as BCD; in this case the curve will consist of two branches, each of unlimited extent. This is the *hyperbola*.

In these three positions of the cutting plane are included two cases, which perhaps deserve separate notice; namely, that in which the cutting plane is perpendicular to the axis of the cone and the section consequently a circle, and that in which the plane passes through the vertex and the section is two straight lines: but these positions of the cutting plane need not be further alluded to, because the circle may be considered as a particular case of an ellipse, and the two straight lines of an hyperbola.

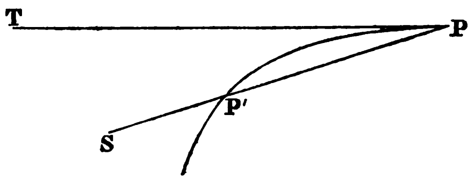
We may say therefore that there are only three different sections of a cone, the *parabola*, the *ellipse*, and the *hyperbola*, and it will be our business to study the properties of these CONIC SECTIONS in order.

It may be remarked, by the way, that the Conic Sections are curves of especial interest for three reasons; first, on account of the simplicity and elegance of their properties; secondly, because of their historical interest as curves known and studied with success by the ancients; and, thirdly, because science has taught us that they are what may be called *physical* curves. A stone when projected describes a parabola, the planets move in ellipses, many comets describe parabolas, some perhaps hyperbolas.

Although we have spoken of the Conic Sections as the sections of a cone, which is a mode of proceeding rendered appropriate by the name usually given to the three curves in question, we shall find it convenient in treating of their properties to adopt other definitions, and we shall have to shew that the curves so defined are really conic sections according to our present notion. It is convenient to conceive of a

curve as traced by a point which moves according to an assigned law; thus we may consider a circle as a curve traced by a point, which moves under the condition of being always at the same distance from a fixed point; and this is the mode of definition which we shall adopt in the case of each of the conic sections; we shall call the curves so defined by the names of the Parabola, Ellipse, and Hyperbola, and afterwards prove that the curves defined are the three sections of a cone.

As we shall have much to do with the tangents to the conic sections, we will here explain the proper notion of the tangent of a curve. Let P be a point in a curve, P' a contiguous point, draw the *secant* $PP'S$, that is the line *cutting* the curve in P and P'. Now suppose P' to approach P, then when

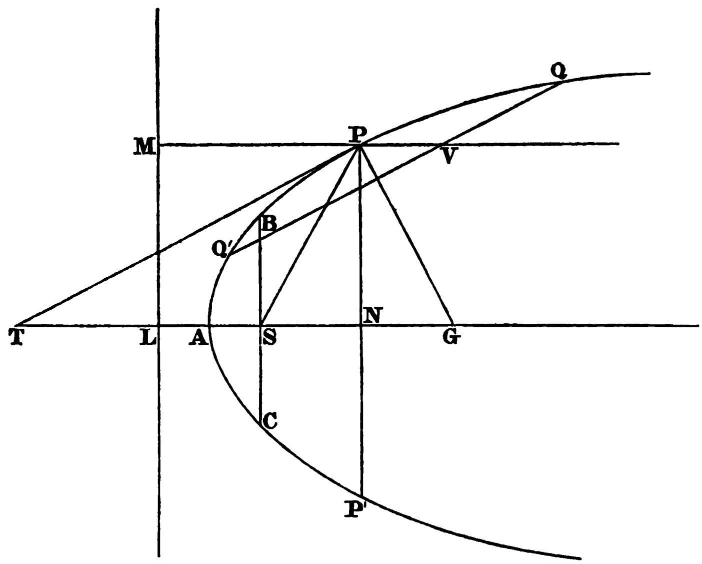


P' and P are *indefinitely* near together, the *secant* $SP'P$ will become the *tangent* TP. In other words, a tangent may be conceived of as a secant, drawn through two points in the curve indefinitely near to each other.

NOTE. It will be understood, that in this subject an algebraical notation is used for the purpose of abbreviation only. Thus $AB \cdot CD$ will be merely a contracted form of the phrase "the rectangle under AB, CD."

Also it may be observed, that in the figures the dotted lines refer wholly to the corollaries of the propositions.

THE PARABOLA.



Dᴇғ. If a point P move in such a manner, that its distance SP from a given point S is always equal to its perpendicular distance PM from a given fixed line LM, the curve traced out by P is called a *parabola*.

The fixed line LM is called the *directrix*, and the point S the *focus*.

Draw SL perpendicular to the directrix and bisect it in the point A: it is manifest, from the definition, that A is a point in the parabola; this point is called the *vertex*.

It is further manifest, that the curve will be exactly similar on opposite sides of the line AS produced: this line is called the *axis*.

Hence it also follows that a line PNP' drawn through P perpendicular to the axis to meet the parabola in P' will be

bisected in N: PNP' is called an *ordinate*, and the line AN an *abscissa* of the axis.

The *ordinate* BC, through the focus, is called the *latus rectum*.

Any line MPV parallel to the axis is called a *diameter*.

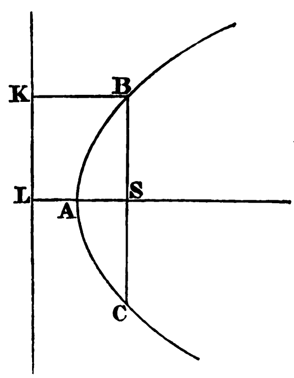
The names *abscissa* and *ordinate* are not confined to lines measured along the axis and perpendicular to it; they are also applied to lines measured along any diameter and parallel to the tangent at the extremity of that diameter: thus if QVQ' be drawn parallel to the tangent PT, PV is called an *abscissa of the diameter*, and QVQ' an *ordinate*. The propriety of this nomenclature will be seen hereafter, when it is proved that the properties of the lines PV, QVQ' are exactly analogous to those of AN, PNP'.

If PT be a tangent to the curve at P, NT is called the *subtangent*.

The *normal* PG is a line perpendicular to the tangent. NG is called the *subnormal*.

PROP. I.

The latus rectum BC = 4AS.



Draw BK perpendicular to the directrix, and produce SA to meet the directrix in L; then

$$SB = BK \text{ by definition,}$$
$$= SL = 2AS, \text{ since } AS = AL \text{ by definition;}$$
$$\therefore BC = 2SB = 4AS.$$

Prop. II.

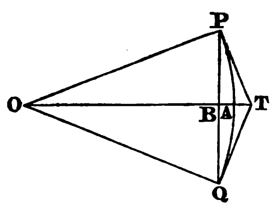
The tangent at any point of a parabola bisects the angle between the focal distance and the diameter through the point.

Let P be a point in the parabola, P' a contiguous point: draw the secant $TP'P$*, join SP, SP', and draw PM, $P'M'$, perpendicular to the directrix, and $P'm$, $P'n$ perpendicular to PM, SP respectively.

Then we shall assume what will hereafter be proved†, namely, that when P and P' are indefinitely near together $P'n$ will coincide with the arc of the circle described with centre S and radius SP', in other words that Sn will be equal to SP'.

* The student is requested to join the points P, P' in the figure.

† Newton, Section I. Lemma VII. This may be also proved trigonometrically: for let PQ be a very small arc of a circle having O for its centre; join PQ, and draw the tangents PT, QT, and let OT intersect the arc PQ and the chord PQ in A and B respectively. Then it is manifest that



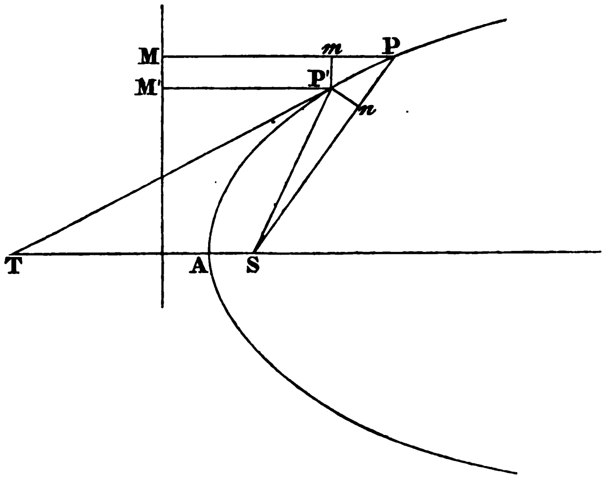
$$PAQ \text{ is } > PBQ \text{ and } < PT + QT,$$

or that PA is $> PB$ and $< PT$,

or that $\dfrac{PA}{OP}$ is $> \sin POA$ and $< \tan POA$,

but $\dfrac{\sin POA}{\tan POA} = \cos POA = 1$ when POA is indefinitely small. Therefore PB and PT are equal when PA is indefinitely small, and therefore PA which is intermediate in value is equal to either of them.

The student will comprehend the force of this measuring more completely, when he has read the first section of Newton's Principia.



Hence $Pn = SP - SP'$

$= PM - P'M'$ (by definition of parabola)

$= Pm$;

therefore in the right-angled triangles PmP', PnP', we have the side $Pm =$ the side Pn, and PP' common; therefore the triangles are equal in all respects, and $\angle\, mPP' = \angle\, nPP'$.

But when P' has approached indefinitely near to P, mPP' is the angle between the tangent and the perpendicular on the directrix, and nPP' is the angle between the tangent and the focal distance. And these angles have been shewn to be equal, therefore the proposition is true.

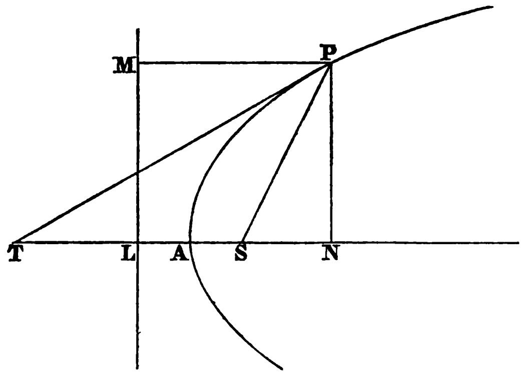
Cor. 1. The tangent at the vertex is perpendicular to the axis.

Cor. 2. The normal bisects the angle between the focal distance and the diameter at the point.

Cor. 3. If T, G (see figure, page 132) are the intersections of the tangent and normal respectively with the axis, $SP = ST = SG$. (Eucl. I. 5.)

Prop. III.

The subtangent is equal to twice the abscissa. (NT=2AN).



Draw PM perpendicular to the directrix LM; then

$$ST = SP = PM = LN,$$

or $AT + AS = AL + AN$;

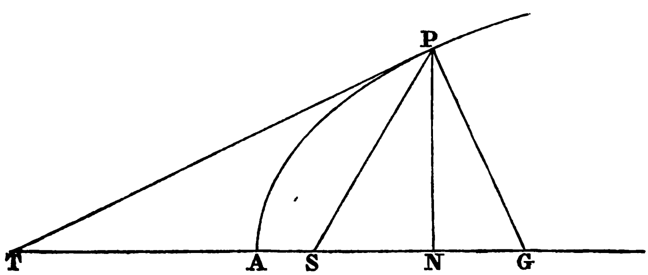
and $AS = AL$;

$\therefore\ AT = AN,$

or $NT = 2AN.$

Prop. IV.

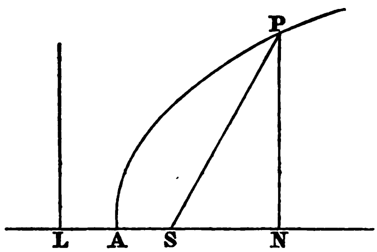
The subnormal is constant, and equal to half the latus rectum. (NG = 2AS).



$$SG = ST$$
$$= AS + AT$$
$$= AS + AN \text{ (by preceding Prop.)}$$
$$= 2AS + SN;$$
$$\therefore \; NG = SG - SN = 2AS.$$

PROP. V.

The rectangle under the latus rectum and the abscissa is equal to the square of a semi-ordinate of the axis. ($PN^2 = 4AS \,.\, AN$).



Because AN is divided into two parts in S,

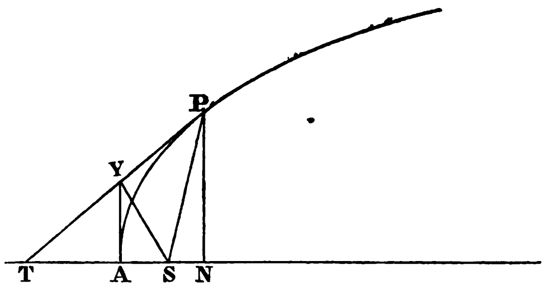
$$\therefore \; 4AS \,.\, AN + SN^2 = LN^2 \quad \text{(Euclid, II. 8.)}$$
$$= SP^2$$
$$= PN^2 + SN^2;$$
$$\therefore \; PN^2 = 4AS \,.\, AN.$$

PROP. VI.

If a perpendicular is drawn from the focus on the tangent, the point of intersection lies in the tangent at the vertex.

Let AY be the tangent at the vertex intersecting the tangent PT in Y; join SY, which shall be perpendicular to PT.

Because AY is parallel to PN, and $AT = AN$; therefore $TY = PY$. (Euclid, VI. 2.)



Also $SP = ST$, and SY is common to the two triangles SPY, STY: therefore these two triangles have their sides respectively equal, and are therefore equal in all respects.

Therefore $\angle SYP = \angle SYT$, and therefore each is a right angle. Hence SY is perpendicular to PT, and therefore the proposition is true.

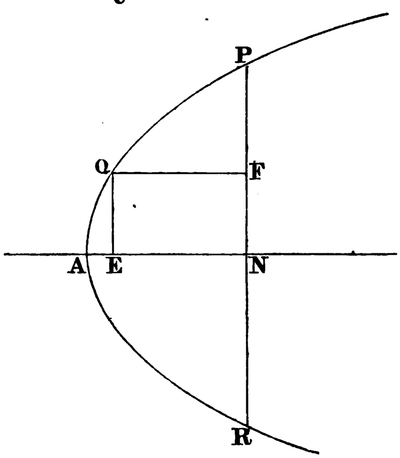
Cor. $SY^2 = SP \cdot AS$.

For from similar triangles SAY, SYP,

$$AS : SY :: SY : SP.$$

Prop. VII.

If from any point F *in the ordinate* PR *the line* FQ *is drawn parallel to the axis and meeting the parabola in* Q, *then* $PF \cdot FR = 4AS \cdot QF$.



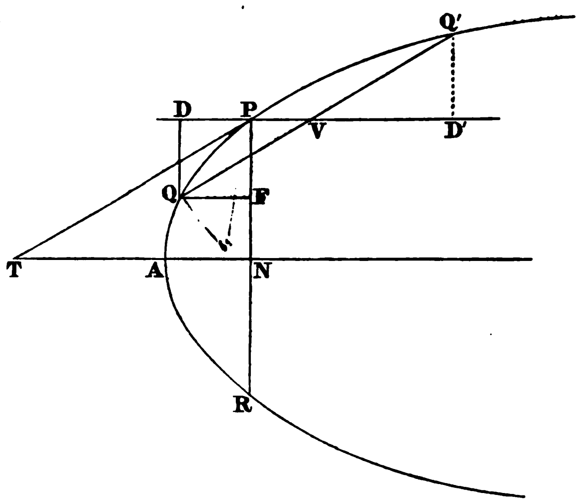
Draw QE perpendicular to the axis; then because PR is divided into two equal parts in N and two unequal in F,

$$\begin{aligned} \therefore \quad PF \cdot FR &= PN^2 - NF^2 \quad \text{(Euclid, II. 5.)} \\ &= PN^2 - QE^2 \\ &= 4AS \cdot AN - 4AS \cdot AE \\ &= 4AS \cdot EN = 4AS \cdot QF. \end{aligned}$$

Prop. VIII.

If from either extremity of an ordinate QVQ′ *a perpendicular* QD *is let fall on the diameter, then*

$$QD^2 = 4AS \cdot PV.$$



Draw the tangent PT, and QF perpendicular to the ordinate PNR, then from similar triangles QDV, PNT,

$$QD : DV :: PN : NT,$$

but $$PN^2 = 4AS \cdot AN = 2AS \cdot NT;$$

$$\begin{aligned} \therefore \quad PN : NT &:: 2AS : PN \\ &:: 4AS : PR; \end{aligned}$$

$$\therefore \quad QD : DV :: 4AS : PR,$$

$$\text{or} \quad 4AS \,.\, DV = QD \,.\, PR = PF \,.\, PR.$$

Also (Prop. VII.)

$$4AS \,.\, QF = PF \,.\, FR,$$

$$\text{or} \quad 4AS \,.\, DP = PF \,.\, FR;$$

$$\therefore \quad 4AS \,.\, PV = PF \,.\, PR - PF \,.\, FR$$

$$= PF^2,$$

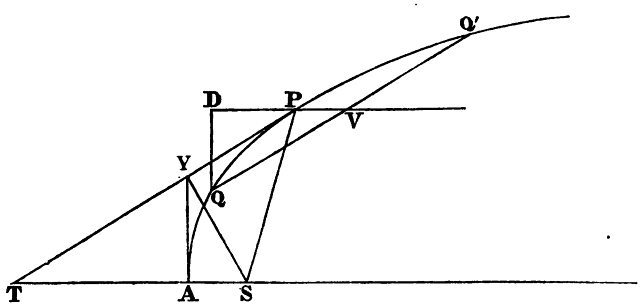
$$\text{or} \quad QD^2 = 4AS \,.\, PV.$$

The proof would be similar if we were to draw $Q'D'$ perpendicular to the diameter from Q'.

COR. $QD = Q'D'$, and therefore $QV = Q'V$, or a diameter bisects all its ordinates.

PROP. IX.

The square of a semi-ordinate of the diameter at any point is equal to four times the rectangle under the focal distance of the point and the abscissa. ($QV^2 = 4SP \,.\, PV$).



Draw AY the tangent at the vertex, SY perpendicular to the tangent PT, QD perpendicular to the diameter, and join SP. Then by similar triangles QDV, SAY,

$$QV^2 : QD^2 :: SY^2 : AS^2,$$

but from similar triangles SAY, SYP,

$$AS : SY :: SY : SP,$$

$$\text{or} \quad SY^2 = AS \, . \, SP;$$

$$\therefore \quad QV^2 : QD^2 :: SP : AS,$$

$$\text{but} \quad QD^2 = 4AS \, . \, PV; \quad \text{(Prop. VIII.)}$$

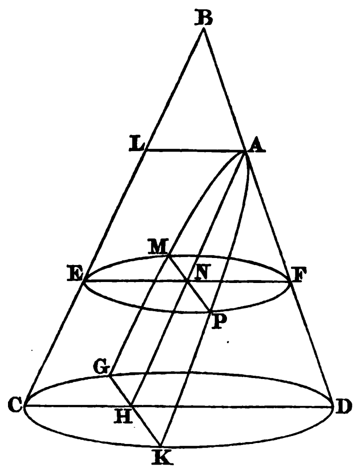
$$\therefore \quad QV^2 : 4AS \, . \, PV :: SP : AS,$$

$$\text{or} \quad QV^2 = 4SP \, . \, PV.$$

Obs. It will be seen that this proposition includes Prop. v., since in that case P coincides with A and $SP = AS$.

Prop. X.

If a right cone is cut by a plane which is parallel to a line in its surface, the section is a parabola.



Let BCD be the section of the cone by the plane of the paper, AGK the cutting plane which is supposed perpendicular to the plane of the paper and parallel to BC. Let $EMFP$ be any circular section made by a plane perpendicular to the axis of the cone. Then the line MNP, in which the planes $EMFP$, AGK intersect, is manifestly perpendicular to both of the lines ENF, ANH, in which those planes intersect the plane of the paper. Draw AL parallel to CD.

Then by similar triangles ANF, BCD,

$$AN : NF :: BC : CD;$$

but the ratio $BC : CD$ is a constant ratio,

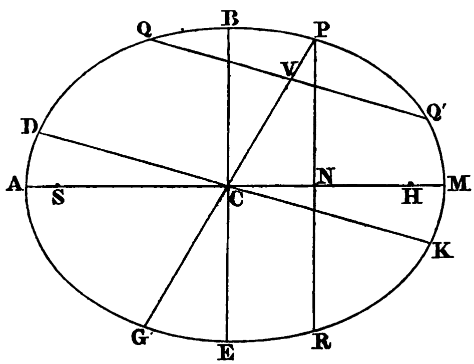
$$\therefore AN \propto NF.$$

Again, since EPF is a semicircle,

$$NF : PN :: PN : EN$$
$$:: PN : AL,$$
$$\text{or} \quad NF \cdot AL = PN^2:$$

but AL is a line of constant length, therefore $NF \propto PN^2$; and we have also proved that $AN \propto NF$, therefore $PN^2 \propto AN$, which is a property of the parabola, (Prop. v.): hence the curve GAK is a parabola.

THE ELLIPSE.



Def. If a point P move in such a manner that the sum of its distances from two fixed points S, H is always the same, the curve traced out by P will be an *ellipse*.

The points S, H are called the *foci*, and the point C bisecting SH the *centre*.

Any line PCG through the centre is called a ***diameter***: it is manifest that the centre bisects all such lines.

The diameter $ASHM$ through the foci is called the *axis major*: A, M are called *vertices*.

A line PNR perpendicular to the axis major is called an *ordinate*, and the lines AN, NM *abscissæ*, of the axis.

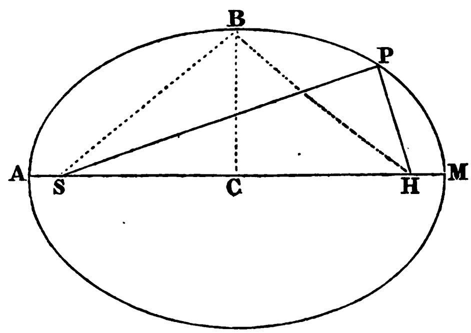
The ordinate BCE through the centre is called the *axis minor*, and that through either focus the *latus rectum*.

The diameter DCK which is parallel to the tangent at P is said to be *conjugate* to PCG.

A line QVQ' parallel to the conjugate diameter is called an *ordinate* of the diameter PCG, and the lines PV, VG *abscissæ*.

Prop. I.

The sum of the focal distances of any point is equal to the axis major. (SP + HP = 2AC).



For, by definition,

$$SP + HP = SA + HA$$
$$SP + HP = SM + HM;$$
$$\therefore\ 2(SP + HP) = SA + SM + HA + HM$$
$$= 2AM;$$
$$\text{or } SP + HP = AM = 2AC.$$

Cor. $SB + HB = 2AC$.

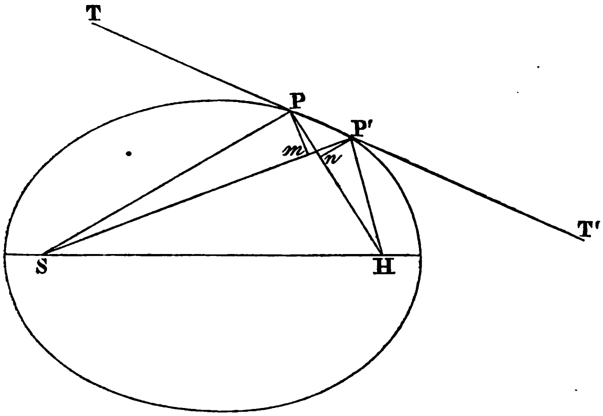
But manifestly $SB = HB$;

$$\therefore\ SB = AC.$$

Hence also $BC^2 = AC^2 - SC^2$.

Prop. II.

The tangent at any point of an ellipse makes equal angles with the focal distances.



Let P, P' be two contiguous points in the ellipse; draw the secant $TPP'T'$; join SP, SP', HP, HP', and draw Pm, $P'n$ perpendicular to SP', HP respectively.

Then (as in the case of the parabola,) when P and P' are indefinitely near to each other, $P'm = SP' - SP$.

In like manner

$$Pn = HP - HP',$$

$$\text{but} \quad SP + HP = SP' + HP';$$

$$\therefore \; P'm = Pn.$$

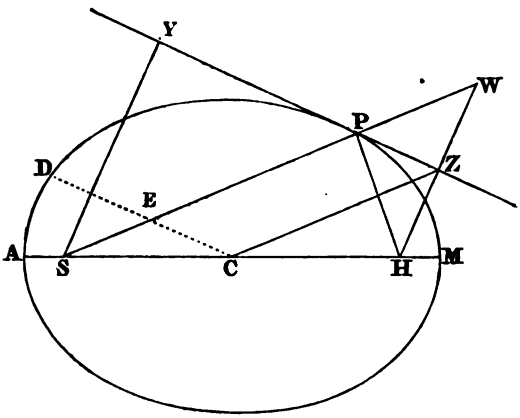
And therefore in the right-angled triangles PmP', PnP', we have the side $P'm = Pn$, and the side PP' common: hence the triangles are equal in all respects, and $\angle PP'm = \angle P'Pn$: but when P and P' coincide these angles are those which the tangent makes with the focal distances: hence the proposition is true.

Cor. 1. The tangent bisects the angle between HP and SP produced.

Cor. 2. The tangent at either vertex is perpendicular to the axis major.

Prop. III.

The perpendiculars from the foci on the tangent intersect the tangent in the circumference of a circle, having the axis major as diameter.



Produce SP to W, making $PW = HP$: join WH, cutting the tangent in Z: join CZ.

Then in the triangles HPZ, WPZ we have the sides HP, WP equal by construction, PZ common, and the angles HPZ, WPZ equal by the property of the tangent: therefore the triangles are equal in all respects, and $\angle PZH = \angle PZW$, each of which is therefore a right angle. Hence HZ is the perpendicular on the tangent.

Again, $SC = CH$, and $WZ = ZH$; $\therefore$ CZ is parallel to SW; and by similar triangles CZH, SWH, $CZ = \frac{1}{2} SW$. But $SW = SP + PW = SP + PH = 2AC$; $\therefore$ $CZ = AC$, and therefore Z is a point in a circle, the centre of which is C and radius AC.

The proof would have been the same, if we had considered SY the perpendicular from the focus S.

Cor. Draw the conjugate diameter CD cutting SP in E. Then $PECZ$ is a parallelogram.

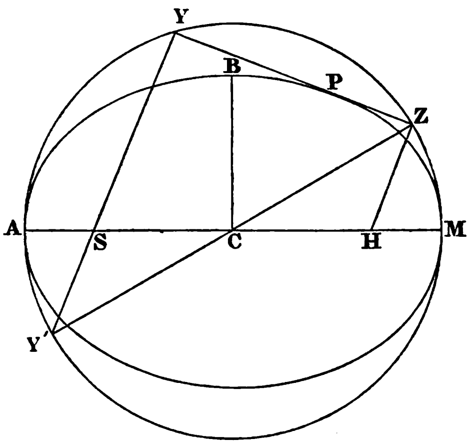
$$\therefore\ PE = CZ = AC.$$

Note. As the circle, which is the subject of the preceding proposition, is of great use in demonstrating the properties of the ellipse, we shall call it the *auxiliary circle*.

Prop. IV.

The rectangle under the perpendiculars from the foci on the tangent is equal to the square of the semi-axis minor.

$$(SY \cdot HZ = BC^2.)$$



Produce SY to meet the auxiliary circle in Y'; join CY', CZ. Then in the triangles SCY', ZCH, we have $SC = CH$, $CY' = CZ$, and $\angle CSY' = \angle CHZ$ since SY' and HZ are parallel; therefore the triangles are equal in all respects, and $SY' = HZ$.

$$\therefore\ SY \cdot HZ = SY \cdot SY' = AS \cdot SM; \quad \text{(Euclid, III. 35.)}$$

$$\text{but } AS \cdot SM = AC^2 - SC^2 \quad \text{(Euclid, II. 5.)}$$

$$= BC^2; \quad \text{(Prop. I. Cor.)}$$

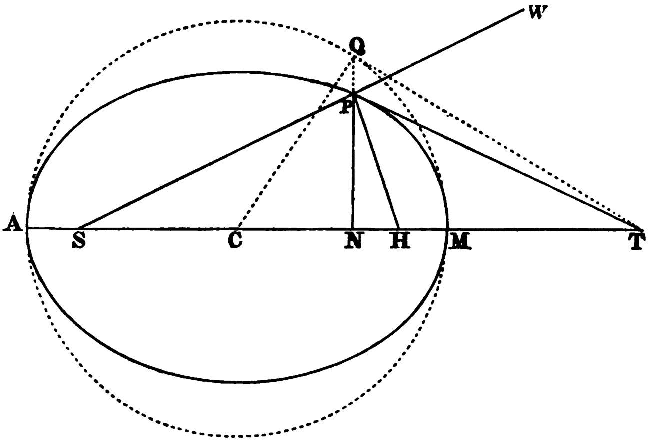
$$\therefore\ SY \cdot HZ = BC^2.$$

Cor. By similar triangles SYP, HZP, (fig. Prop. III.)

$$SY : HZ :: SP : HP;$$
$$\therefore SY^2 : SY . HZ :: SP : HP,$$
$$\therefore SY^2 : BC^2 :: SP : 2AC - SP.$$

Prop. V.

The rectangle under the lines intercepted between the centre and the intersections of the axis with the ordinate and tangent respectively, is equal to the square of the semi-axis major. (CN . CT = AC².)



Produce SP to W, then because PT bisects the exterior angle WPH,

$$\therefore ST : HT :: SP : HP. \text{ (Euclid, VI. A.)}$$
$$\therefore ST + HT : ST - HT :: SP + HP : SP - HP,$$
$$\text{or } 2CT : SH :: 2AC : SP - HP.$$

Again, we have

$$SP^2 = SN^2 + PN^2,$$
$$HP^2 = HN^2 + PN^2;$$
$$\therefore SP^2 - HP^2 = SN^2 - HN^2,$$

$$\text{or } (SP - HP)(SP + HP)$$
$$= (SN - HN)(SN + HN), \text{ (Eucl. II. 6.)}$$
$$\text{or } (SP - HP)\, 2AC = 2CN \,.\, SH,$$
$$\text{or } 2CN : SP - HP :: 2AC : SH.$$

Hence $CT : AC :: AC : CN$,

$$\text{or } CT \,.\, CN = AC^2.$$

Cor. Produce PN to meet the auxiliary circle in Q. Join CQ, TQ.

Then $CT \,.\, CN = AC^2 = CQ^2$,

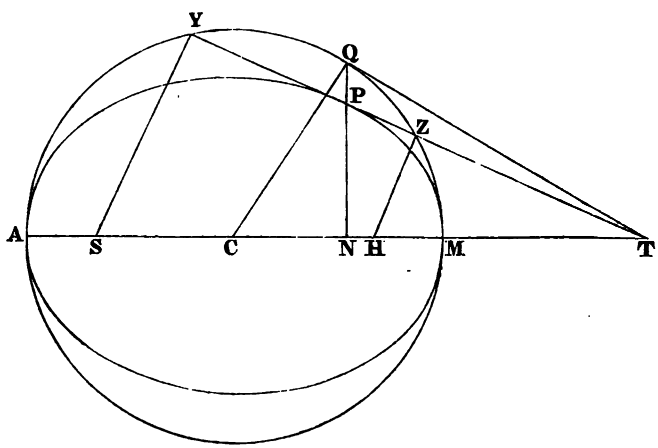
$$\text{or } CT : CQ :: CQ : CN;$$

therefore CQT is a right angle, and QT touches the circle at Q; that is, *the tangents of the ellipse and circle at P and Q respectively cut the major axis produced in the same point.*

Prop. VI.

The rectangle under the abscissæ of the axis major is to the square of the semi-ordinate, as the square of the axis major to the square of the axis minor.

$$(AN \,.\, NM : PN^2 :: AC^2 : BC^2).$$



Draw the tangent PT and the perpendiculars upon it from the foci SY, HZ; produce PN to meet the auxiliary circle in Q, join CQ, and draw the tangent QT: then the triangles SYT, PNT, HZT are similar,

$$\therefore\ PN : SY :: NT : YT,$$

$$\text{and } PN : HZ :: NT : ZT;$$

$$\therefore\ PN^2 : SY.HZ :: NT^2 : YT.ZT,$$

$$\text{or } PN^2 : BC^2 :: NT^2 : QT^2 \quad \text{(Euclid, III. 36.)},$$

$$:: QN^2 : CQ^2 \text{ by similar triangles,}$$

$$:: AN.NM : CQ^2 \text{ by property of the circle,}$$

$$\text{or } AN.NM : PN^2 :: AC^2 : BC^2.$$

COR. Hence if L be the latus rectum,

$$L.AC = 2BC^2.$$

For by the proposition

$$AS.SM : \frac{L^2}{4} :: AC^2 : BC^2,$$

$$\text{or} \quad BC^2 : \frac{L^2}{4} :: AC^2 : BC^2, \text{ (See Prop. IV.)}$$

$$\text{or} \quad BC : \frac{L}{2} :: AC : BC,$$

$$\therefore\ L.AC = 2BC^2.$$

PROP. VII.

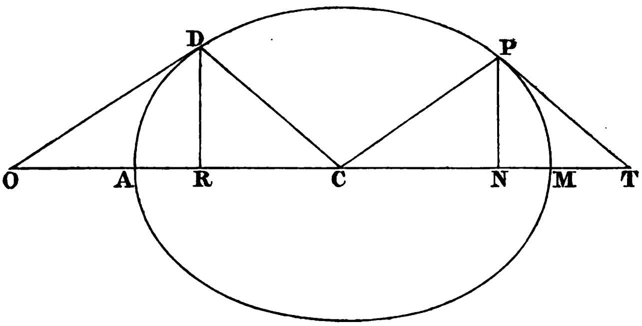
If the semidiameter CD *is conjugate to* CP, *then* CP *is conjugate to* CD.

Draw the ordinates PN, DR and the tangents PT, DO. Then the triangles PNT, CDR are similar.

But $\quad CN.CT = AC^2$;

$$\therefore\ CN.NT = AC^2 - CN^2$$

$$= AN.NM; \quad \text{(Euclid, II. 5.)}$$



in like manner

$$CR \cdot RO = AR \cdot RM;$$

$$\therefore\ CN \cdot NT : CR \cdot RO :: AN \cdot NM : AR \cdot RM,$$
$$:: PN^2 : DR^2,$$
$$:: NT^2 : CR^2,$$
$$\therefore\ CN : RO :: NT : CR,$$
$$:: PN : DR;$$

therefore the triangles PCN, DOR are similar, and CP is parallel to DO, or CP is conjugate to CD.

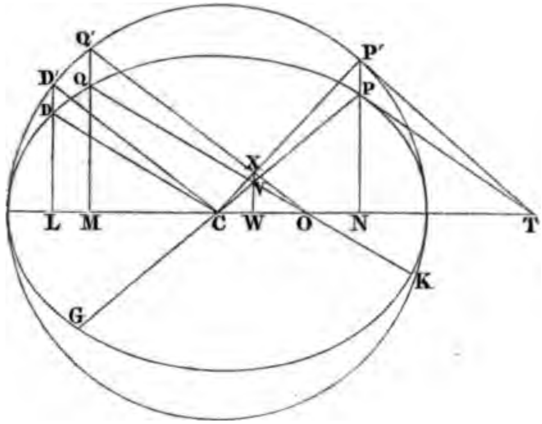
Obs. It is evident that the major and minor axes are *conjugate diameters*.

Prop. VIII.

The rectangle under the abscissæ of any diameter is to the square of the semiordinate, as the square of the diameter to the square of the conjugate.

$$(PV \cdot VG : QV^2 :: CP^2 : CD^2).$$

Let the ordinate QVK meet the axis in O. Produce the ordinates NP, LD to meet the auxiliary circle in P', D' respectively: draw QM, VW perpendicular to the axis: let VW cut CP' in X: join OX, and produce it to meet MQ produced in Q'.



Then $Q'M : QM :: XW : VW$

$:: P'N : PN$;

therefore Q' is a point in the auxiliary circle.

Again, draw the tangents PT, $P'T$ to the ellipse and circle, then CD is parallel to PT.

$$\therefore CL : NT :: DL : PN :: D'L : P'N;$$

therefore the triangles CLD', TNP' are similar, and CD' is parallel to $P'T$ and perpendicular to CP'.

In like manner $Q'X$ is perpendicular to CP'.

$$\text{Now} \quad CP^2 : CV^2 :: CP'^2 : CX^2;$$

$$\therefore CP^2 - CV^2 : CP'^2 - CX^2 :: CP^2 : CP'^2,$$

$$\text{or} \quad PV.VG : Q'X^2 :: CP^2 : AC^2;$$

$$\text{and} \quad Q'X^2 : QV^2 :: CD'^2 : CD^2 :: AC^2 : CD^2;$$

$$\therefore PV.VG : QV^2 :: CP^2 : CD^2.$$

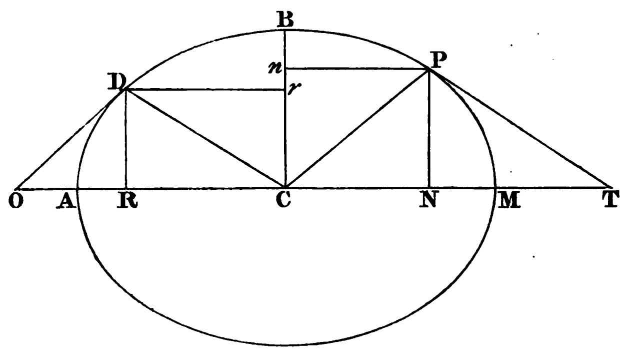
Obs. It will be seen that Prop. vi. is only a particular case of this proposition, since in that case $CP = AC$ and $CD = BC$. It has already been remarked that the axes of the ellipse are conjugate diameters.

Prop. IX.

√ *The sum of the squares of conjugate diameters is constant.*

$$(CP^2 + CD^2 = AC^2 + BC^2).$$

Draw the tangents PT, DO, the ordinates PN, DR, and Pn, Dr ordinates to the minor axis.



Then $CN \cdot CT = AC^2 = CR \cdot CO$;

$\therefore\ CN : CR :: CO : CT$,

$:: CD : PT$ by similar triangles CDO, CPT;

$:: CR : NT$ by similar triangles CDR, TPN;

or $CR^2 = CN \cdot NT = CN \cdot CT - CN^2$

$= AC^2 - CN^2$;

or $CR^2 + CN^2 = AC^2$.

Similarly $Cn^2 + Cr^2 = BC^2$;

$\therefore\ CN^2 + Cn^2 + CR^2 + Cr^2 = AC^2 + BC^2$,

or $CP^2 + CD^2 = AC^2 + BC^2$.

Cor. Since $CR^2 = AC^2 - CN^2$, $\therefore$ CR = the ordinate in the auxiliary circle corresponding to the point P;

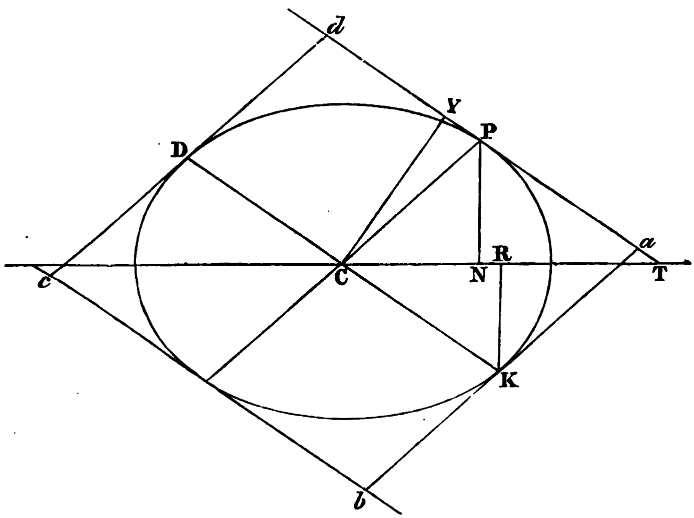
$\therefore\ PN : CR :: BC : AC$;

similarly, $DR : CN :: BC : AC$.

Prop. X.

Parallelograms circumscribing an ellipse, the sides of which are parallel to conjugate diameters, have the same area.

Let $abcd$ be the parallelogram formed by tangents parallel to the semiconjugate diameters CP, CD. Draw CY perpendicular to the tangent PT, and the ordinates PN, KR.



Then by similar triangles CYT, CKR,

$$CT : CY :: CK : KR;$$

$$\therefore CY . CK = CT . KR;$$

$$\text{but } CT : AC :: AC : CN,$$

$$\text{and } KR : CN :: BC : AC \quad \text{(Prop. IX. Cor.)}$$

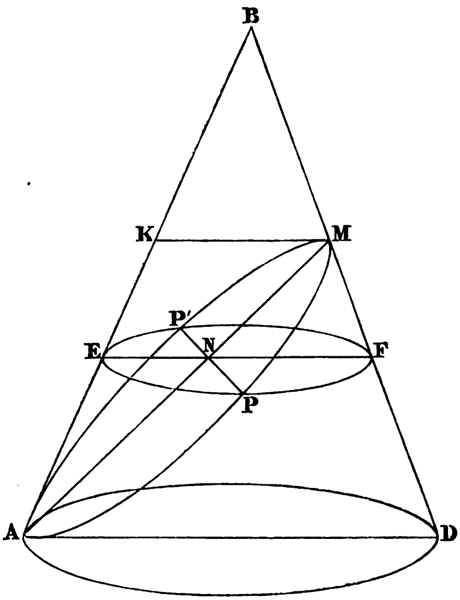
$$\therefore CT . KR = AC . BC,$$

$$\text{and } \therefore CY . CK = AC . BC;$$

but $CY . CK$ is one fourth of the parallelogram $abcd$, therefore the parallelogram $abcd = 4AC.BC$ = the rectangle under the axes of the ellipse.

Prop. XI.

If a right cone is cut by a plane which is not parallel to a line in its surface, and the section is wholly on one side of the vertex, the section is an ellipse.



Let BAD be the section of the cone by the plane of the paper, $APMP'$ the cutting plane which is supposed perpendicular to the plane of the paper. Let $EPFP'$ be any circular section made by a plane perpendicular to the axis of the cone. Then the line PNP', in which the planes $EPFP'$, $APMP'$ intersect, is manifestly perpendicular to both EF and AM. Draw MK parallel to AD. Then by similar triangles,

$$AN : EN :: AM : KM,$$

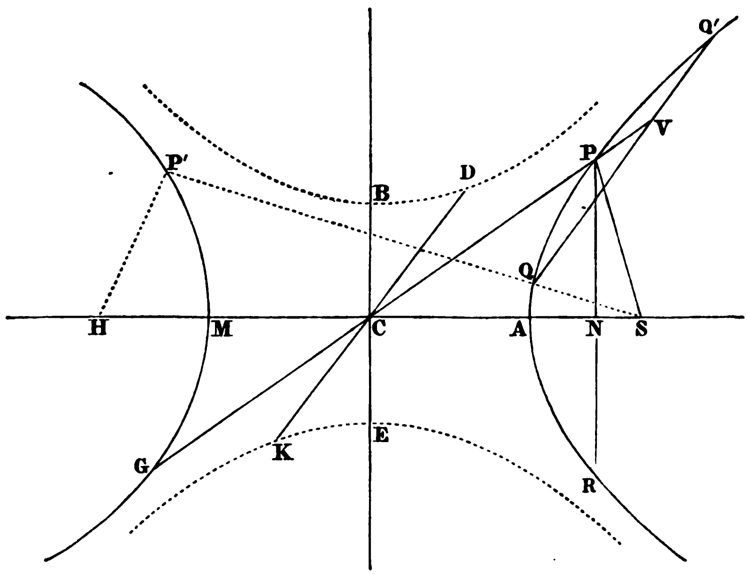
$$\text{and } NM : NF :: AM : AD;$$

$$\therefore AN . NM : EN . NF :: AM^2 : KM . AD,$$

$$\text{or } AN . NM : PN^2 :: AM^2 : KM . AD;$$

but this is a property of an ellipse, the minor axis of which is a mean proportional between KM and AD; hence the section is an ellipse.

THE HYPERBOLA.



Def. If a point P move in such a manner that the difference of its distances from two fixed points S, H is always the same, the curve traced out by P will be an *hyperbola.*

The points S, H are called the *foci*, and the point C bisecting SH the *centre.*

It is evident that the hyperbola must consist of *two branches*, because for every point P to the right of C there will be another P' to the left of it, exactly similarly situated; consequently the curve will be exactly symmetrical with respect to C. Moreover, it is clear that the branches will be infinite, since there is no limit put to the magnitudes of HP and SP by the condition of their *difference* being constant.

The definition of the hyperbola being so nearly analogous to that of the ellipse, it will be anticipated that many of their

properties will be possessed in common; also the nomenclature is so devised, as to preserve as much as possible the analogy between the two curves.

Any line PCG through the centre and meeting the two branches of the curve is called a *diameter*: it is manifest that the centre bisects all such lines.

The diameter ACM, which produced passes through the foci, is called the *axis major*: A, M are called *vertices*.

PNR is an *ordinate* of the axis; AN, NM the *abscissæ*.

A line BC drawn from C, perpendicular to the axis major, and such that $BC^2 = SC^2 - AC^2$ is called the *semi-axis minor*; $BE (= 2BC)$ is the *axis minor*; this is according to the analogy of the ellipse, in which we had $BC^2 = AC^2 - SC^2$. (Ellipse, Prop. I. Cor.)

An hyperbola (represented by a dotted line in the figure) having BE for its major and AM for its minor axis, is called the *conjugate hyperbola*.

When the major and minor axes are equal, the hyperbola is said to be *rectangular*; it will be seen that the rectangular has the same kind of relation to the common hyperbola, as the ellipse has to the circle.

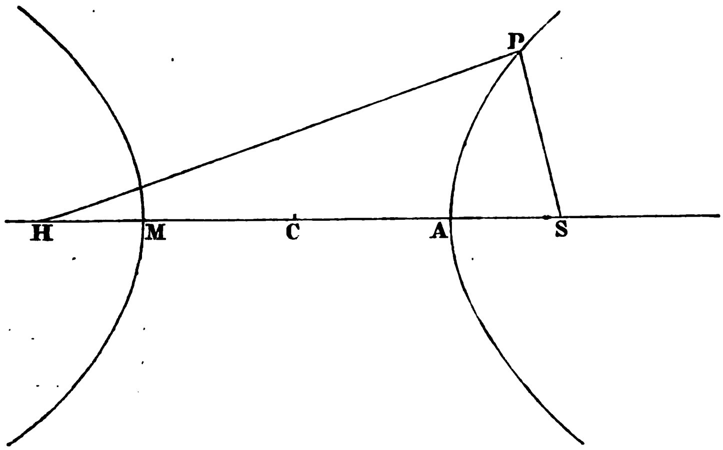
A line KCD drawn through C, parallel to the tangent at P, to meet the conjugate hyperbola, is called the *diameter conjugate* to PCG.

Other definitions are the same as those for the ellipse.

An *asymptote* to a curve is a line which continually approaches the same, but being produced ever so far does not cut it.

Prop. I.

The difference of the focal distances of any point is equal to the axis major. (HP − SP = 2AC).



For, by definition,

$$HP - SP = HA - SA,$$

$$HP - SP = SM - HM;$$

$$\therefore\ 2(HP - SP) = HA - HM + SM - SA,$$

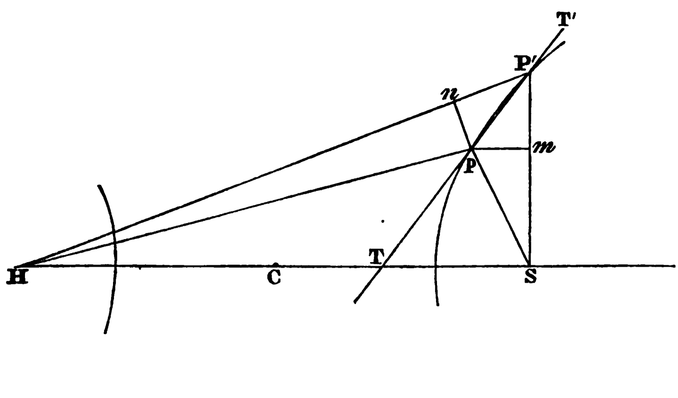
$$= 2AM,$$

$$\text{or } HP - SP = AM = 2AC.$$

Prop. II.

The tangent at any point of an hyperbola makes equal angles with the focal distances.

Let P, P' be two contiguous points in the hyperbola; draw the secant $TPP'T'$; join SP, SP', HP, HP', and draw Pm, Pn perpendicular to SP', HP' respectively.



Then (as in the case of the parabola,) when P and P' are indefinitely near to each other, $P'm = SP' - SP$.

In like manner

$$P'n = HP' - HP;$$
$$\text{but } HP - SP = HP' - SP',$$
$$\therefore\ P'm = P'n,$$

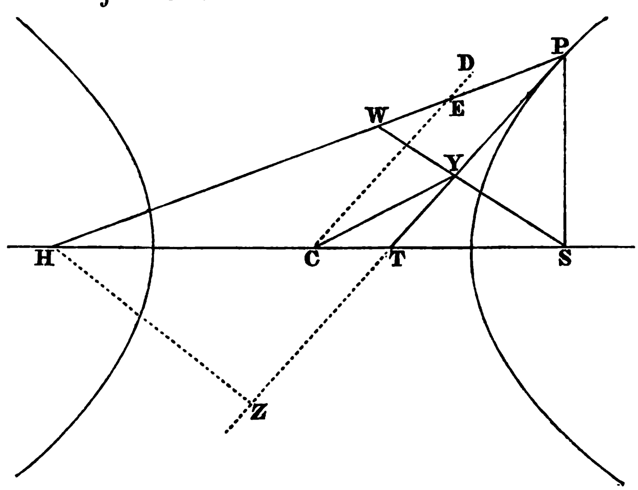
and therefore in the right-angled triangles PmP', PnP', we have the side $P'm = P'n$, and the side PP' common: hence the triangles are equal in all respects, and $\angle PP'm = \angle PP'n$: but when P and P' coincide these angles are those which the tangent makes with the focal distances; hence the proposition is true.

Cor. The tangent at either vertex is perpendicular to the axis major.

Prop. III.

The perpendiculars from the foci on the tangent intersect the tangent in the circumference of a circle having the axis major as diameter.

In HP take $PW = SP$; join SW, cutting the tangent PT in Y: join CY.



Then in the triangles PSY, WPY, we have the sides SP, PW equal by construction, PY common, and the angles SPY, WPY equal by the property of the tangent; therefore the triangles are equal in all respects, and $\angle SYP = \angle WYP$, each of which is therefore a right angle. Hence SY is the perpendicular on the tangent.

Again, $SC = CH$ and $SY = YW$, therefore CY is parallel to HW; and by similar triangles CSY, HSW, $CY = \frac{1}{2}HW$.

But $HW = HP - PW = HP - SP = 2AC$.

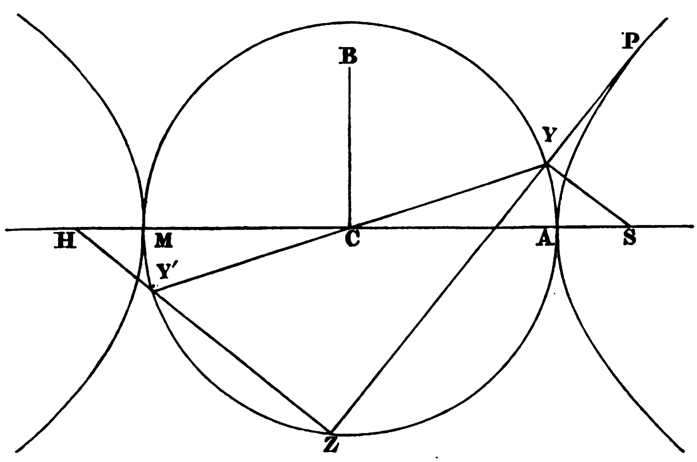
Therefore $CY = AC$, and therefore Y is a point in a circle the centre of which is C and radius AC.

The proof would have been the same if we had considered HZ the perpendicular from H, or if we had drawn the tangent to the other branch of the curve.

Cor. Draw the semi-conjugate CD cutting HP in E: then $CYPE$ is a parallelogram, and $PE = CY = AC$.

Prop. IV.

The rectangle under the perpendiculars from the foci on the tangent is equal to the square of the semi-axis minor. $(SY \cdot HZ = BC^2)$.



Let the perpendicular HZ meet the auxiliary circle also in Y'. Join CY, CY'; then, in the triangles CSY, CHY', we have $CS = CH$, $CY = CY'$, and $\angle CSY = \angle CHY'$ since SY, HY' are parallel; therefore the triangles are equal in all respects, and $SY = HY'$,

$$\begin{aligned} \therefore\ SY.HZ &= HY'.HZ \\ &= HM.HA \\ &= HC^2 - CM^2 \quad \text{(Eucl. II. 6.)} \\ &= BC^2 \text{ by definition.} \end{aligned}$$

Cor. By similar triangles SYP, HZP, (fig. Prop. III.)

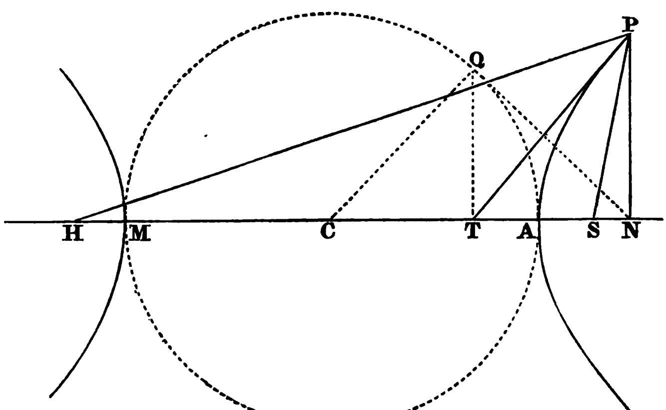
$$SY : HZ :: SP : HP,$$

$$\therefore\ SY^2 : SY.HZ :: SP : HP,$$

$$\therefore\ SY^2 : BC^2 :: SP : 2AC + SP.$$

Prop. V.

The rectangle under the lines intercepted between the centre and the intersections of the axis with the ordinate and tangent respectively is equal to the square of the semi-axis major. ($CN . CT = AC^2$).



Because PT bisects the angle HPS,

$$\therefore\ HT : ST :: HP : SP, \text{ (Euc. III. 6)};$$

$$\therefore\ HT - ST : HT + ST :: HP - SP : HP + SP,$$

$$\text{or } 2CT : SH :: 2AC : HP + SP.$$

Again, we have

$$SP^2 = SN^2 + PN^2,$$

$$HP^2 = HN^2 + PN^2;$$

$$\therefore\ HP^2 - SP^2 = HN^2 - SN^2,$$

$$\text{or } (HP + SP)(HP - SP) = (HN + SN)(HN - SN), \quad \text{(Euc. II. 6)}$$

$$\text{or } (HP + SP)2AC = 2CN . SH,$$

$$\text{or } 2CN : HP + SP :: 2AC : SH.$$

Hence $$CT : AC :: AC : CN,$$

$$\text{or } CT . CN = AC^2.$$

COR. Draw TQ an ordinate to the auxiliary circle, and join NQ, CQ.

Then $$CT . CN = AC^2 = CQ^2,$$

$$\text{or } CT : CQ :: CQ : CN.$$

Therefore CQN is a right angle, and NQ *touches* the circle.

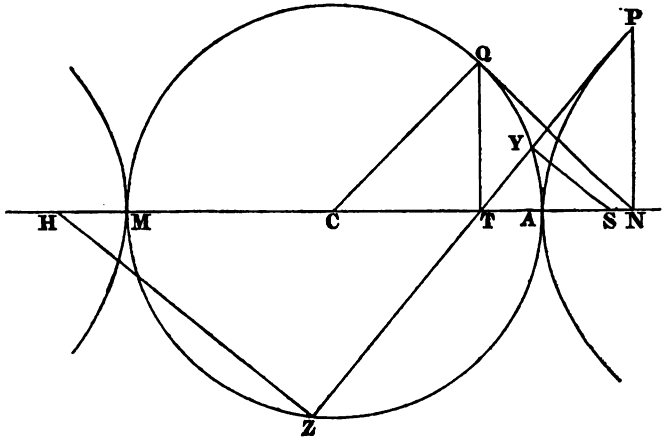
PROP. VI.

The rectangle under the abscissæ of the axis major is to the square of the semiordinate, as the square of the axis major to the square of the axis minor.

$$(AN . NM : PN^2 :: AC^2 : BC^2.)$$

Draw the tangent PT and the perpendiculars upon it from the foci SY, HZ; draw TQ an ordinate to the auxiliary

circle, and join NQ, CQ. Then the triangles SYT, PNT, HZT are similar;



$$\therefore PN : SY :: NT : YT,$$

$$\text{and } PN : HZ :: NT : ZT;$$

$$\therefore PN^2 : SY.HZ :: NT^2 : YT.ZT,$$

$$\text{or } PN^2 : BC^2 :: NT^2 : QT^2, \text{ (Euc. III. 35.)}$$

$$:: QN^2 : CQ^2, \text{ by similar triangles,}$$

$$:: AN.NM : AC^2, \text{ (Euc. III. 36.)}$$

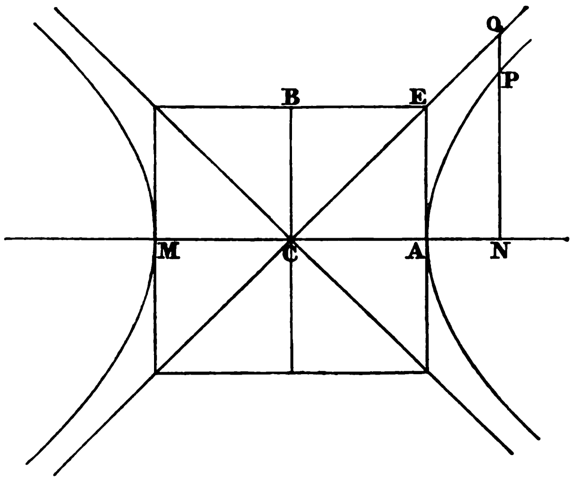
$$\text{or } AN.NM : PN^2 :: AC^2 : BC^2.$$

Cor. As in the case of the ellipse,

$$L . AC = 2BC^2.$$

Prop. VII.

If tangents be drawn at the vertices of the hyperbola and of the conjugate hyperbola, the diagonals of the rectangle so formed will be asymptotes to the hyperbola.



Let CE be a diagonal of the rectangle, and let the ordinate PN produced meet it in Q.

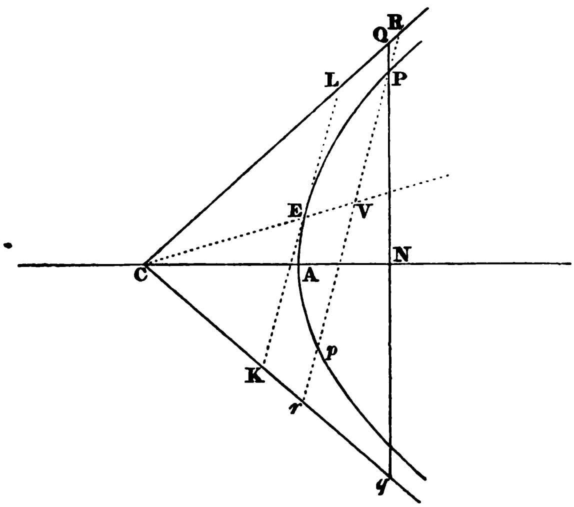
Then $$NQ^2 : CN^2 :: AE^2 : AC^2$$
$$:: BC^2 : AC^2$$
$$:: PN^2 : AN.NM$$
$$:: PN^2 : CN^2 - AC^2.$$

Now as CN increases, the ratio of $CN^2 : CN^2 - AC^2$ approaches to equality, but never actually attains to it. Hence also NQ never actually becomes equal to PN, but continually approximates to it in value, and therefore the line CQ is an asymptote to the curve.

Cor. The same line is an asymptote to the conjugate hyperbola.

Prop. VIII.

If any line perpendicular to the axis is terminated by the asymptotes, the rectangle under the segments into which it is divided by the curve is equal to the square of the semi-axis minor. $(QP.Pq = BC^2)$.



Let QPq be the line terminated by the asymptotes.

Then

$$NQ^2 : PN^2 :: CN^2 : CN^2 - AC^2; \text{ (Prop. VII.)}$$

$$\therefore NQ^2 - PN^2 : PN^2 :: AC^2 : CN^2 - AC^2$$

$$\text{or } NQ^2 - PN^2 : AC^2 :: BC^2 : AC^2,$$

$$\therefore NQ^2 - PN^2 = QP \,.\, Pq = BC^2.$$

COR. 1. If any line RPr terminated by the asymptotes makes a given angle with either asymptote, the rectangle $RP \,.\, I'r$ is invariable.

$$\text{For } RP : QP :: \sin PQR : \sin PRQ,$$

$$\text{and } Pr : Pq :: \sin Pqr : \sin Prq;$$

$$\therefore RP.Pr : QP.Pq\,(BC^2) :: \sin PQR.\sin Pqr : \sin PRQ.\sin Prq.$$

Now all the angles in this proportion are the same so long as RPr makes a given angle with either asymptote, and therefore $RP \,.\, Pr$ is invariable.

Cor. 2. If the line RPr cuts the curve also in p, we have, manifestly,

$$RP \cdot Pr = Rp \cdot pr,$$

$$\text{or } RP \cdot Pp + RP \cdot pr = RP \cdot pr + Pp \cdot pr;$$

$$\therefore RP = pr.$$

Cor. 3. The same conclusions will hold, if we suppose RPr to move parallel to itself until it becomes a tangent LEK. Hence we have $LE = EK$, and $RP \cdot Pr = LE^2$.

Cor. 4. Join CE, and let it when produced cut Pp in V. Then,

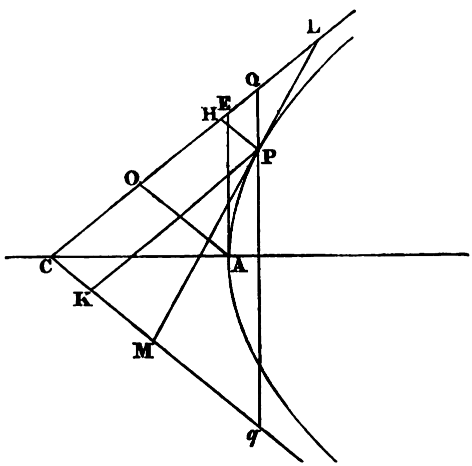
$$VR : LE :: VC : EC$$

$$:: Vr : EK;$$

but $LE = EK$, $\therefore VR = Vr$; also $RP = rp$; therefore $PV = Vp$, or a diameter bisects its ordinates.

Prop. IX.

If from any point in the curve straight lines are drawn parallel to and terminated by the asymptotes, their rectangle is invariable.



Let PH, PK be the lines parallel to the asymptotes. Draw the tangent LPM, and QPq perpendicular to the axis,

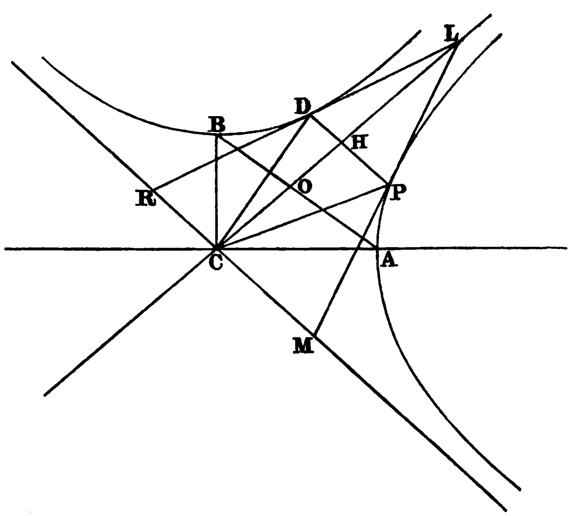
AE the tangent at the vertex, and AO parallel to the asymptote.

Then, by similar triangles,

$$PH : PQ :: AO : AE,$$
$$PK : Pq :: OE : AE;$$
$$\therefore PH.PK : PQ.Pq\ (BC^2) :: AO.OE\ (AO^2) : AE^2\ (BC^2),$$
$$\text{or } PH.PK = AO^2$$
$$= \tfrac{1}{4}(AC^2 + BC^2).$$

Cor. 1. The parallelogram $PHCK$ is constant; and therefore also the triangle LCM is constant, for since $PL = PM$ it is double of the parallelogram $PHCK$.

Cor. 2. A straight line drawn parallel to one asymptote and terminated by the conjugate hyperbolas, is bisected by the other asymptote.



Let PHD be such a line, then by the proposition

$$PH.HC = \tfrac{1}{4}(AC^2 + BC^2),$$

similarly for the conjugate hyperbola,

$$DH \,.\, HC = \tfrac{1}{4}(AC^2 + BC^2)\,;$$

$$\therefore\; PH = DH.$$

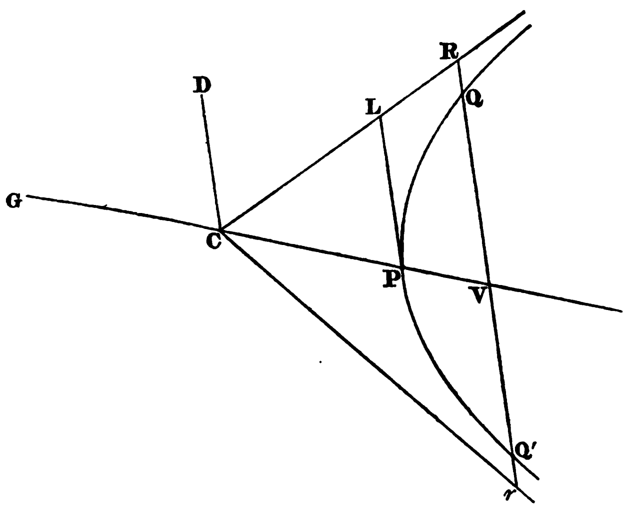
Cor. 3. If we draw the tangent LPM, it is bisected in P, and therefore $CL = 2CH$; hence the tangent RDL to the conjugate hyperbola at D, must meet the asymptote in the same point L. Also $CM = 2HP = DP$, therefore $CDPM$ is a parallelogram, and CD is the semidiameter conjugate to CP.

Cor. 4. $RL = 2CP =$ the diameter at P; and $LM = 2CD$ = the diameter conjugate to CP.

Prop. X.

The rectangle under the abscissæ of any diameter is to the square of the semiordinate, as the square of the diameter to the square of the conjugate.

$$(PV \,.\, VG : QV^2 :: CP^2 : CD^2).$$



Let QVQ' produced meet the asymptotes in R, r; and

draw PL a tangent at P terminated by the asymptote, then PL equals the semiconjugate CD.

Now $$RQ \,.\, Qr = RV^2 - QV^2 = PL^2,$$

$$\therefore\ QV^2 = RV^2 - PL^2;$$

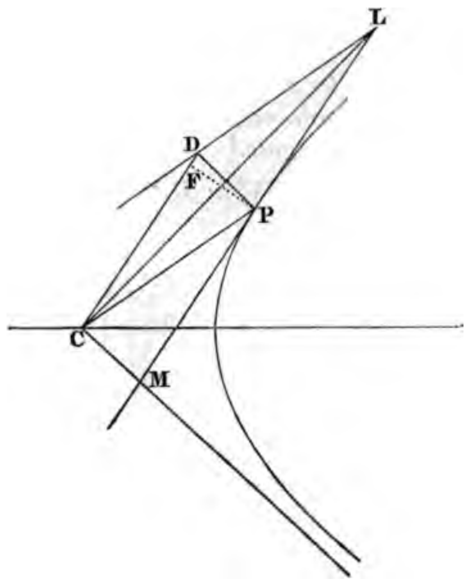
$$\text{but } CV^2 : CP^2 :: RV^2 : PL^2:$$

$$\therefore\ CV^2 - CP^2 : RV^2 - PL^2 :: CP^2 : PL^2,$$

$$\text{or } PV \,.\, VG : QV^2 : CP^2 : CD^2.$$

PROP. XI.

The parallelograms formed by tangents at the vertices of pairs of conjugate diameters have all the same area.



Let MPL, DL be the tangents, CL, CM the asymptotes; then the parallelogram in this case

$$= 4CDLP$$

$$= 8CLP = 4LCM,$$

which is constant, since LCM is constant. (Prop. IX. Cor. 1.)

Cor. Draw PF perpendicular to CD, then

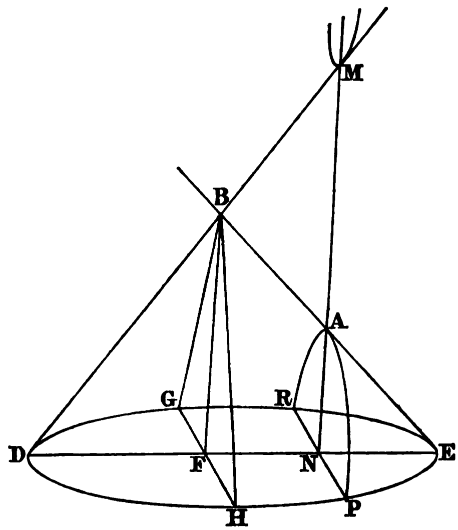
$$PF \cdot CD = CDLP;$$

but when the tangents are drawn at A and B, the area of the parallelogram $CDLP$ becomes $AC \cdot BC$, therefore

$$PF \cdot CD = AC \cdot BC.$$

Prop. XII.

If a right cone is cut by a plane which meets the cone on both sides of the vertex, the section is an hyperbola.



Let RAP be the section, which is supposed to be perpendicular to the plane of the paper; $DGEH$ a section of the cone perpendicular to the axis, which is therefore circular, RNP the intersection of the planes RAP, $DGEH$, which is manifestly perpendicular to both DE and AN. BGH a triangular section through the vertex of the cone by a plane parallel to RAP.

Then $$AN : EN :: BF : EF,$$

$$NM : ND :: BF : FD,$$

$$\therefore\ AN \cdot NM : EN \cdot ND :: BF^2 : EF \cdot FD,$$

$$\text{or } AN \cdot NM : PN^2 :: BF^2 : FH^2,$$

which is the property of an hyperbola, the major axis of which is AM, and the minor axis is to AM as $FH : BF$; hence the section is an hyperbola.

NOTE. For some other propositions concerning the Conic Sections, see the Digression concerning the curvature of curves in the First Section of Newton's Principia.

SCHOLIUM.

Although, throughout this treatise, the symbol $AB \cdot CD$ has been used to express the rectangle under the two lines AB, CD, and not to express as in algebra the multiplication of AB by CD, nevertheless the symbol may be regarded in this algebraical point of view: for if a be the number of units of length in AB, and b the number in CD, then will $a \times b$ be the number of units of area in the rectangle $AB \cdot CD$, and therefore $AB \cdot CD$ may be regarded as a product; it being understood that in so regarding it, AB represents the number of units of length in the line AB, and CD the number in the line CD. In like manner the ratio $AB : CD$ may be written thus $\frac{AB}{CD}$, with the same understanding. The importance of this scholium will be appreciated by the student, when he sees the application of the properties of the Conic Sections in the sequel.

MECHANICS.

I. STATICS.

II. DYNAMICS.

MECHANICS.

The science of *mechanics* treats of the effects of *force*; we must therefore commence by explaining what we mean by *force*.

Force is any cause which changes, or tends to change, a body's state of rest or motion.

By the term *body* here used, we intend to express any material substance, or portion of matter; we cannot conceive of the action of a force except as taking place upon a material body; and the body may be of various kinds, but at present we shall concern ourselves only with the action of force upon a *particle*, and the action of force upon a *rigid body*. By a *particle* we intend, without entering into any discussion respecting the ultimate constitution of matter, to designate the smallest quantity of matter conceivable, so that we need not concern ourselves with its shape or its magnitude; nevertheless all particles are not necessarily equal, for though each of two particles be indefinitely small, one may be greater than the other in any proportion. By a *rigid body* we mean to denote any portion of matter, the constituent particles of which are so connected as to be incapable of changing their relative position. From this definition it is easy to see, that no such thing as a mathematically rigid body exists in nature, for the hardest known substances are susceptible of compression and extension under the action of great pressures; steel, for instance, though a very hard substance, is not rigid; nevertheless the conclusions which we shall arrive at in the following treatise, though mathematically true only of rigid bodies, will be practically true of all ordinary solid bodies: for, to consider a body as absolutely rigid, or incapable of changing its form under the action of any force, is the same thing practically as to suppose, that no forces are called into

play, which actually do produce any sensible change of form in the solid body under consideration.

When any number of forces act upon a material body they will produce one of two effects, they will either keep the body at rest or they will cause motion, and these two effects are of such very distinct kinds that they require to be treated separately; and thus the science of mechanics naturally divides itself into two parts, the first and more simple of which treats of forces which keep a body at *rest*, or are in *equilibrium*, and is called *Statics;* the second of forces which produce *motion*, and is called *Dynamics*.

STATICS.

1. Force is measured statically by the pressure which will counteract it. So far as the principle of measuring force is concerned, it is indifferent what kind of pressure we choose as the standard by which to measure force; there is one kind of pressure, however, which in the nature of things is more convenient than any other, and that is *weight*. We may conceive then of force as being measured by the number of pounds which it can lift, and we may compare two forces by means of the numbers of pounds which they can lift respectively.

In what follows we shall denote the *magnitude* or *intensity* of force by letters, such as P, or Q, or R; when we speak of *a force* P, all that is intended is that the force in question would just lift P lbs., and therefore has P lbs., or more shortly P, for its statical measure.

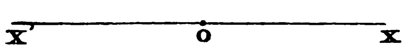
2. In order that we may be able to calculate the combined effect of any number of forces acting on a particle, it is not sufficient that we should know the intensity of each, for it is obvious that the effect of a force depends upon the *direction* in which it acts as well as its magnitude, and that a system of forces cannot in general be in equilibrium unless their directions as well as their magnitudes satisfy certain conditions, which conditions the science of Statics must teach us. Let us endeavour to form a distinct conception of the *direction* of a force: suppose a force to act upon a particle at rest, and the particle to be prevented from moving by a string, one end of which is fixed and the other attached to the particle, then the direction of the force coincides with the string; or we may say, that the direction of a force is that, in which the particle would begin to move, if not constrained to remain at rest. The direction of a force is sometimes called the *line of its action*.

3. The *intensity* and *direction* are the only elements necessary to entirely describe any force which acts on a *particle*, but if we consider its action on a rigid body we shall require in addition to know the *point of its application*: at present we shall concern ourselves only with the action on a single particle.

The simplest case of such action is when there are two forces only, and it is manifest that, in this case, there can be equilibrium only when the two forces are equal in intensity and exactly opposite in their direction: that is, if two forces P and Q are in equilibrium, they must act in the same line and tend to draw the particle opposite ways, and we must have the condition,

$$P = Q, \text{ or } P - Q = 0.$$

4. We may here observe, that the method, which was adopted in the Treatise on Trigonometry, of denoting *direction* by an algebraical sign may be applied to forces: that is, if we denote by $+P$ a force acting on a particle at O in the direction OX, then a force of the same magnitude, but acting in the opposite direction OX' will be properly denoted by $-P$. Hence we may say, that if any number of forces act along the same straight line on a particle, the condition of equilibrium is, that their *algebraical* sum shall be zero; and if the sum be not zero, then the force represented by it will be the *resultant* of the forces acting on the particle, and will tend to draw the particle in the direction OX, or in the direction OX', according as its sign is $+$ or $-$.



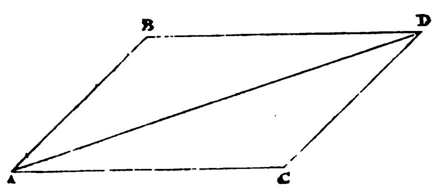
For instance, suppose we have two forces P, Q acting on a particle at O in the direction OX, and two others R, S in the direction OX', then the resultant will be $P + Q - R - S$, the algebraical sum of the four forces, and in order that there may be equilibrium, we must have,

$$P + Q - R - S = 0.$$

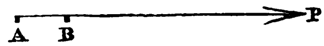
5. When two forces act on a particle, not along the same straight line, they cannot be in equilibrium, but will be equivalent to some one force which we shall call their *resultant:* that they are equivalent to some one force is manifest from the consideration, that a particle under the action of two forces would *begin* to move in a certain definite direction, and that it may be prevented from moving by a string attached to it and coinciding with that direction; the string spoken of will undergo a certain *tension*, and that tension or the weight which would produce it measures the magnitude of the resultant of the two forces.

6. The fundamental problem of Statics is, to find the magnitude and direction of the resultant of two forces; but in order to solve it, we must premise that it is convenient to represent *forces* by *lines;* and it is manifest that we may by means of a line represent a force, as to both magnitude and direction; for we can represent the magnitude, by taking a line which bears the same proportion to some standard length, (as for instance 1 inch,) as the force bears to the standard pressure, (or 1 lb.); and the direction will be represented, by drawing the line in that direction in which the force tends to make the particle move. By means of this convenient mode of representing forces, we are able to enunciate the relation between two forces and their resultant in the form of the following Theorem, which is known as the PARALLELOGRAM OF FORCES.

If two forces acting on a particle at A, *be represented in direction and magnitude by the lines* AB, AC, *then the resultant will be represented in direction and magnitude by the diagonal* AD *of the parallelogram described upon* AB, AC.



7. The proof which we shall give of this proposition depends upon this principle; *a force may be supposed to act at any point in its direction, that point being conceived to be rigidly attached to the particle on which the force acts.* Thus, if we have a force P acting on a particle at A, we may suppose, if we please, that the force acts at B, B being rigidly connected with A; this is a principle which the student will have no difficulty in grasping, and which may be illustrated roughly by saying, that the force required to toll a bell is independent of the length of the rope, and the effort required to move a carriage independent of the length of the traces. We are now able to give the following.



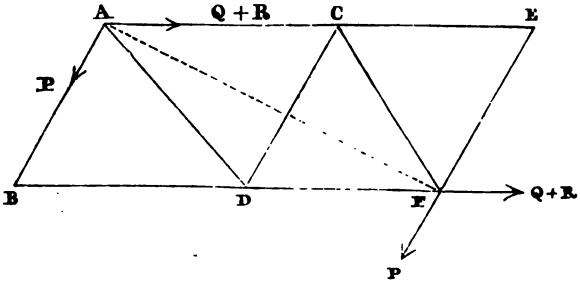
PROOF OF THE PARALLELOGRAM OF FORCES.

8. I. To prove the proposition so far as the *direction* of the resultant is concerned.

When the forces are *equal*, it is manifest that the direction of the resultant will bisect the angle between the directions of the forces: or, if we represent the forces in direction and magnitude by two lines drawn from the point at which they act, the diagonal of the parallelogram described upon these lines will be the direction of the resultant.

Next, suppose that the proposition just proved for equal forces, is true for two unequal forces P and Q, and also for P and R: we shall shew that it will be true for P and $Q + R$.

Let A be the point of application of the forces; take AB to represent P in direction and magnitude, and AC to represent Q; complete the parallelogram $ABDC$, then by hypothesis AD is the direction of the resultant of P and Q; and, since a force may be supposed to act at any point of its direction, we may consider D as the point of application of the



resultant of P and Q, or we may suppose the forces P and Q themselves to act at D, P parallel to AB and Q to AC, or still further, we may suppose P to act at C in the direction CD.

Again, the force R which acts at A may be supposed to act at C; take CE to represent it in direction and magnitude, and complete the parallelogram $CDFE$, then by hypothesis CF is the direction of the resultant of P and R which acted at C: hence the resultant of P and R may be supposed to act at F, or P and R may themselves be supposed to act at that point parallel to their original directions.

Lastly, the force Q, which is at present supposed to be acting at D in the direction DF, may be supposed to act at F.

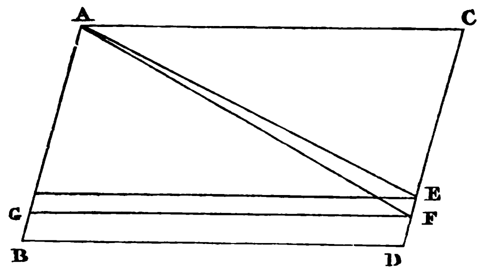
Hence we have reduced the forces P and $Q + R$ acting at A, to P and $Q + R$ acting in the same directions at F; consequently F is a point in the line of action of the resultant, and therefore AF is the direction of the resultant: that is, if the proposition be true for P and Q, and also for P and R, it is true for P and $Q + R$.

But the proposition is true for P and P, and also for P and P, therefore it is true for P and $2P$, therefore for P and $3P$, and so on, therefore generally for P and mP.

In like manner the proposition may be extended to mP and nP, (m and n being whole numbers,) that is, to any *commensurable* forces.

The proposition is extended to incommensurable forces as follows.

Let AB, AC represent any two incommensurable forces; complete the parallelogram $ABDC$, and if AD is not the

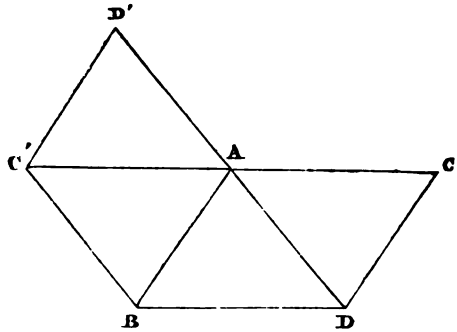


direction of the resultant, let it be AE. Suppose AC to be divided into a number of equal parts, each part being less than ED, and suppose distances of the same magnitude to be set off along CD beginning at C, then one of the divisions must fall between E and D; let F be the point which marks the division, and complete the parallelogram $AGFC$, then AF is the direction of the resultant of the *commensurable* forces AG, AC: but AF makes a larger angle with AC than AE, that is, the resultant of AG and AC lies further away from AC than the resultant of AB and AC, although AG is less than AB, which is absurd: hence AE is not the direction of the resultant, and it may be shewn in like manner that no other line is in that direction except AD. Hence the proposition, which was proved for *commensurable* forces, is true for *incommensurable*.

II. To prove the parallelogram of forces with respect to the *magnitude* of the resultant.

Let AB, AC represent the forces; complete the parallelogram $ABDC$, join DA and produce it to D', making AD' *equal* to the resultant of AB and AC *in magnitude;* complete the parallelogram $ABC'D'$, and join AC'.

Then since AD' is equal to the resultant of AB and AC, and drawn in the direction opposite to that of their resultant, the



three forces AB, AC, AD' will balance each other, and therefore any one of them is in the direction of the resultant of the other two; hence AC is in the direction of the resultant of AB, AD'; but AC' is also in that direction, therefore AC, AC' are in the same straight line. Hence $ADBC'$ is a parallelogram; therefore $AD = BC'$: but $BC' = AD'$, therefore $AD = AD'$. And by construction AD' represents the resultant of AB and AC in magnitude, therefore AD also represents the resultant, and the proposition enunciated is true.

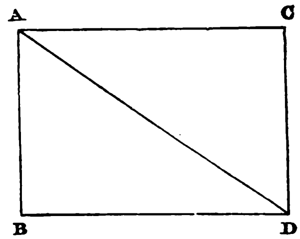
9. The proposition, which we have now established, enables us to reduce any system of forces acting on a particle to one single force; for we can find the resultant of any two of the forces, then of that resultant and a third, and so on.

10. As the parallelogram of forces enables us to compound two forces into one, so conversely we are able by means of it to *resolve* any force into two, that is, to find two forces which shall be equivalent to a given force. This is a problem which obviously admits of an infinite number of solutions; in fact, if upon a line, representing the given force in direction and magnitude, as diagonal we describe any parallelogram, the sides of that parallelogram will represent forces, which by their composition are equivalent to the given force.

Before proceeding further, we shall supply the student with a few examples of the composition and resolution of forces.

Ex. 1. Two forces, measured by 3 lbs. and 4 lbs. respectively, act on a particle at right angles to each other; find the magnitude of their resultant.

If AB, AC represent the forces, and complete the rectangle $ABDC$, we have



$$AD^2 = AB^2 + AC^2$$

$$= 3^2 + 4^2$$

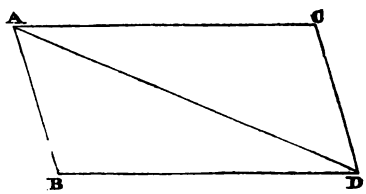
$$= 9 + 16 = 25;$$

$$\therefore\ AD = 5,$$

or the measure of the resultant is 5 lbs.

Ex. 2. Two forces, 1 and 2 lbs. respectively, act at an angle of 60°; find the direction and magnitude of their resultant.

Let AB and AC represent the forces, AD their resultant, and let $BAD = \theta$. Then by the data of the problem, $BAC = 60^\circ$.



$$\therefore\ AD^2 = AC^2 + CD^2 - 2AC \,.\, CD \cos ACD$$

$$= AC^2 + AB^2 + 2AC \,.\, AB \cos BAC$$

$$= 4 + 1 + 4 \cos 60^\circ$$

$$= 5 + 2 \quad (\text{since } \cos 60^\circ = \tfrac{1}{2})$$

$$= 7;$$

$$\therefore\ AD = \sqrt{7}.$$

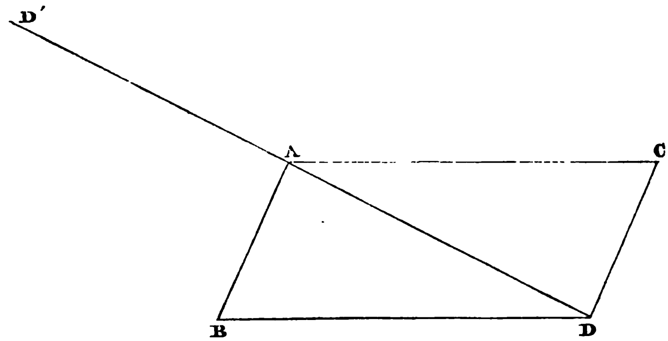
Again, from the triangle ABD,

$$\frac{\sin\theta}{\sin ABD} = \frac{BD}{AD},$$

$$\text{or,}\quad \sin\theta = \frac{2}{\sqrt{7}} \sin 60^\circ = \sqrt{\tfrac{3}{7}}.$$

Ex. 3. Three forces P, Q, R are in equilibrium; find the angles between their directions.

Let AB, AC represent P and Q respectively; complete the parallelogram $ABDC$, and produce DA to D', making $AD' = AD$; then AD' represents R.



Let $BAC = \theta$; $\therefore$ $ACD = 180^\circ - \theta$,

and $AD^2 = AC^2 + CD^2 + 2AC \, . \, CD \cos\theta$;

$$\therefore \quad \cos\theta = \frac{AD^2 - AC^2 - CD^2}{2AC \, . \, CD}$$

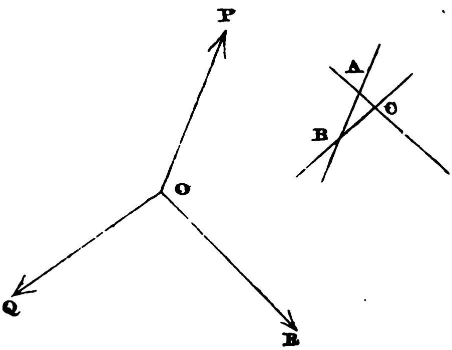
$$= \frac{R^2 - P^2 - Q^2}{2PQ},$$

which determines the angle between the directions of P and Q. The angles between the directions of P and R, Q and R, are known in like manner.

11. The parallelogram of forces may be stated in another form, under which it is called the *Triangle of Forces*. For we have seen that three forces will be in equilibrium, provided they are proportional to the sides and diagonal of a parallelogram, and act on a particle parallel to those sides; but the sides and diagonal form a triangle; indeed, it is the same thing whether we say of three straight lines that they are the sides and diagonal of a parallelogram, or that they will

form a triangle; hence we may assert, that forces will be in equilibrium, when they are proportional to the sides of a triangle formed by drawing lines parallel to their directions.

Suppose, in fact, that the forces P, Q, R are in equilibrium on the particle O; draw any three straight lines paral-



lel to the directions of the three forces, and let ABC be the triangle formed by their intersections, then

$$P : Q : R :: AB : BC : AC.$$

Cor. If α, β, γ are the angles between the directions of Q and R, R and P, P and Q respectively, then

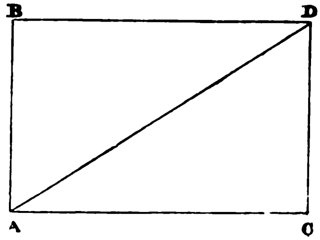
$$P : Q : R :: \sin\alpha : \sin\beta : \sin\gamma;$$

or, (as it may be otherwise written,)

$$\frac{P}{\sin\alpha} = \frac{Q}{\sin\beta} = \frac{R}{\sin\gamma}.$$

12. It follows from the triangle of forces, that any conclusions established concerning the relations of the sides and angles of a triangle, may be extended to the magnitudes and directions of forces in equilibrium. For instance, we may conclude from Euclid I. 19, that of three forces in equilibrium any two must be greater than the third.

13. We have seen that a force in a given direction may always be replaced by two forces in two other directions, which forces are called the *components* of the original force. There is a peculiarity in the case of these components being at right angles to each other, which requires notice. Let AD represent a force, which is resolved into the two rectangular components AB, AC. Then it is manifest, that the force AB has no tendency to move the particle in the direction AC, neither has AC any tendency to move it in the direction AB. Hence we may say that AB, AC measure the *whole* effect of AD in the directions AB, AC respectively, and they are usually termed the *resolved parts* of AD.



If we call the angle BAD θ, we have

$$AB = AD \cos\theta,$$

$$AC = AD \sin\theta.$$

Hence, if X be the *resolved part* of a force P in a direction making an angle θ with the direction of P, and Y the resolved part in the direction perpendicular to that of X, we shall have

$$X = P\cos\theta,$$

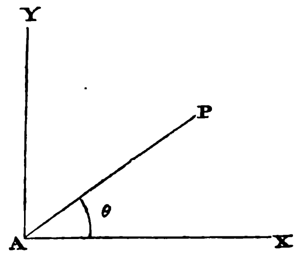
$$Y = P\sin\theta.$$

Also, $\tan\theta = \dfrac{Y}{X}$, and $X^2 + Y^2 = P^2$.

The preceding formulæ may be looked upon as fundamental in Statics; they enable us to solve the following most general problem.

14. *Any number of forces act at the same point, their directions all lying in the same plane; find the direction and magnitude of their resultant.*

Let P be any one of the forces acting at the point A. Let the plane of the paper be that in which the forces act; in that plane choose any two lines at right angles to each other, AX and AY, and let θ be the angle which the direction of P makes with AX. Then P is equivalent to



$P\cos\theta$ acting in the direction AX,

together with $P\sin\theta$ .... AY.

In like manner, a force P', the direction of which makes an angle θ' with AX, is equivalent to

$P'\cos\theta'$ acting in the direction AX,

together with $P'\sin\theta'$ .... AY.

And so on of any number of forces. Hence, adding together the forces which act in the same direction, we shall have a system of forces PP' ... acting at angles $\theta\,\theta'$ ... with the line AX, equivalent to

$P\cos\theta + P'\cos\theta' + \ldots$ acting in the direction AX,

together with $P\sin\theta + P'\sin\theta' +$ .... AY.

For shortness sake, let

$$P\cos\theta + P'\cos\theta' + \ldots\ldots = X,$$

$$P\sin\theta + P'\sin\theta' + \ldots\ldots = Y,$$

and let R be the required resultant, ϕ the angle which its direction makes with the line AX; then

$$R\cos\phi = X,$$

$$R\sin\phi = Y;$$

$$\therefore \tan\phi = \frac{Y}{X}, \quad R^2 = X^2 + Y^2.$$

These formulæ determine the direction and magnitude of the resultant of the system of forces.

15. *To find the conditions of equilibrium of any system of forces, acting in one plane at the same point.*

Suppose the forces are reduced to one (R), as in the last article; then in order that there may be equilibrium we must have

$$R = 0,$$

$$\text{or} \quad X^2 + Y^2 = 0.$$

And this equation is equivalent to these two,

$$X = 0, \qquad Y = 0,$$

$$\text{or} \qquad P \cos\theta + P' \cos\theta' + \ldots\ldots = 0,$$

$$P \sin\theta + P' \sin\theta' + \ldots\ldots = 0.$$

These are the conditions of equilibrium, which may be expressed in words by saying, that *the sum of the forces resolved in any two directions perpendicular to each other must vanish.*

ON THE PRINCIPLE OF THE LEVER.

16. Hitherto we have considered forces acting on a particle only; when we come to the consideration of the action of forces on a rigid body, there will be other conditions of equilibrium besides those already deduced. In the case of a single particle the only necessary condition is that there shall be no motion of *translation*, but in order that a rigid body may be at rest it is not sufficient that any one point in it should be fixed, it is also necessary that there should be no *twisting* about that point. The simplest case is that of two forces acting on a rigid rod, one point of which is fixed; supposing that the forces tend to twist the rod in opposite ways, we can find the conditions, under which they will counteract each the effect of the other, and produce equilibrium.

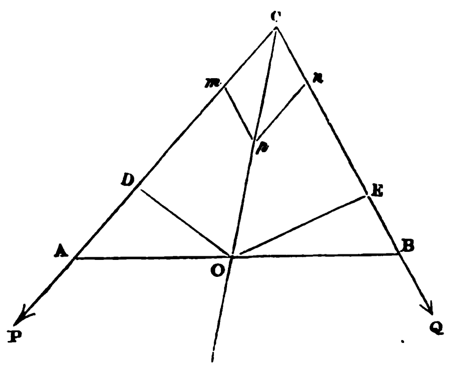
DEF. A rigid rod, moveable about a fixed point in its length, is called a *lever*.

DEF. The fixed point is called the *fulcrum*, and the distances between the fulcrum and the extremities the *arms*.

Def. The *moment* of a force with respect to a given point, is the product of the force and the perpendicular from the point on its direction.

17. Prop. *Two forces acting at the extremities of a lever will produce equilibrium, when the moments of the forces about the fulcrum are equal.*

Let P, Q be the two forces acting at A and B, the extremities of a lever. Produce the directions of P and Q



until they meet in C, then P and Q may both be supposed to act at C: take Cm, Cn proportional to P and Q, and complete the parallelogram $Cmpn$; join Cp and produce it to cut AB in O, then the resultant of P and Q acts in the direction CO, and therefore O must be the fulcrum, otherwise there could not be equilibrium. Draw OD, OE perpendicular to AC, BC; then

$$\frac{P}{Q} = \frac{Cm}{Cn} = \frac{\sin Cpm}{\sin mCp} = \frac{CO \sin BCO}{CO \sin ACO} = \frac{OE}{OD};$$

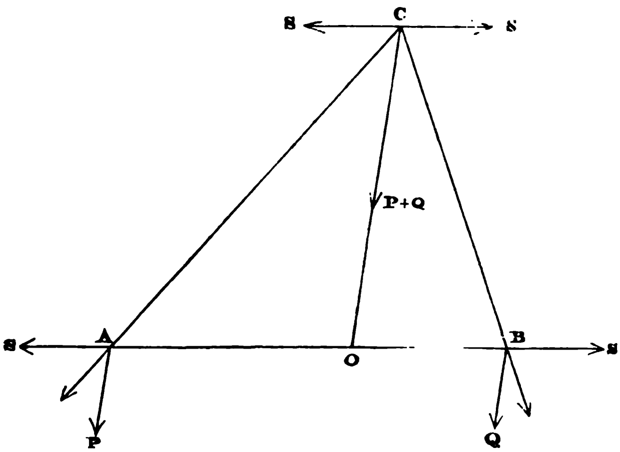
$$\therefore\ P \,.\, OD = Q \,.\, OE,$$

or the moments of P and Q about C are equal.

If the forces act parallel to each other, their directions will not meet as they are supposed to do in the preceding

proposition; in this case we must proceed as follows. We shall suppose, though it is not necessary, that the forces act perpendicular to the lever.

At A and B apply any two equal and opposite forces S in the direction of the lever; this will manifestly not affect



the equilibrium; then the resultant of P and S will be some force in the direction CA suppose, and that of Q and S some force in the direction CB. Suppose them both to act at C, and there to be resolved into their constituent parts P and S, Q and S; the portion S, S will destroy each other, leaving a resultant $P + Q$ in the direction CO parallel to the directions of the forces.

Then the sides of the triangle AOC are parallel to the directions of the forces P, S and their resultant;

$$\therefore \frac{P}{S} = \frac{CO}{AO},$$

in like manner, $$\frac{Q}{S} = \frac{CO}{BO};$$

$$\therefore P \cdot AO = Q \cdot BO.$$

The same method is applicable, when the forces are not perpendicular to the arm.

Hence we conclude, that in all cases two forces, acting in the same plane, on a rigid body, one point of which is fixed, will produce equilibrium when their moments about the fixed point are equal; for it is manifest, that the reasoning applied to the case of the simple lever, is applicable to a rigid body of any shape. This is called the *principle of the lever.*

Ex. 1. Two weights, of 3 and 4 lbs. respectively, balance on the extremities of a lever, the length of which is 6 feet: find the fulcrum.

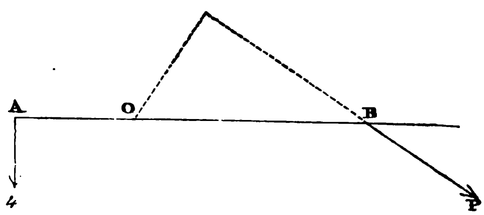
Let x be its distance from that extremity at which the weight of 3 lbs. is suspended; then $6-x$ is its distance from the other;

$$\therefore\ 3x = 4(6-x),$$

$$7x = 24,$$

$$x = \frac{24}{7} = 3\tfrac{3}{7} \text{ feet.}$$

Ex. 2. A weight of 4 lbs. is suspended from a straight lever, at a distance of 2 feet from the fulcrum; determine the force, which acting at an angle of 30^0 with the lever, and at a distance of 3 feet from the fulcrum, will produce equilibrium.



Let P be the required force.

Then we have

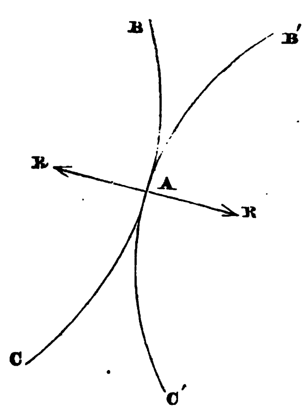
$$P \times 3 \sin 30^0 = 4 \times 2,$$

$$P = \frac{2}{3} \times 8 = \frac{16}{3} = 5\tfrac{1}{3} \text{ lbs.}$$

ON THE ACTION AND REACTION OF SMOOTH SURFACES IN CONTACT.

18. When two bodies are pressed together by the action of any forces, each will exert upon the other a certain force; if we call the force exerted by one body its *action*, we may call that of the other upon the first the *reaction*, and it is evident, upon consideration, that these must be equal in intensity and opposite in direction. For if a person presses his finger upon the table it is manifest that the table returns a pressure equal to, because it is the effect of, the pressure of the finger; and so in other cases.

But what will be the direction of this mutual pressure? We shall consider only the case of the action of two bodies smooth and lying all in one plane. Let BAC, $B'AC'$ be two such bodies, touching at A, and let R be the mutual pressure. Then since the surfaces touch at A, they have a common tangent at that point, and therefore a common normal. Now by calling a body *smooth*, we mean to assert there is no tendency in the constitution of the body to prevent motion along its surface, consequently no part of the mutual pressure of two smooth surfaces can be along the surface or along the common tangent, that is, *the whole must be in the direction of the common normal.*



In the case of a particle pressing upon a curve, we must consider that the pressure is in the direction of the normal to the curve; and when resting on a plane, the pressure is perpendicular to the plane.

We may observe here, that in treating statical problems which involve more than one body, we usually consider what

action and reaction will exist between the various bodies, and having denoted them by certain symbols, as P, Q, R, &c., we consider the equilibrium of each body separately. An example of this will be found in the problem of the wedge (Art. 33, page 200).

ON THE MECHANICAL POWERS.

19. The Mechanical Powers are the elementary forms of all machines, and may be considered as simple devices for enabling a smaller force usually called *the Power* (P), to be in equilibrium with a larger force, usually called *the Weight* (W). They may be thus enumerated:—the Lever, the Wheel and Axle, the Toothed Wheel, the Pully, the Inclined Plane, the Wedge, and the Screw.

(1) *The Lever.*

20. We have already considered the principle of the Lever as a general mechanical principle, and we have shewn that two forces will balance about a fulcrum when their moments about it are equal; but the lever may also be regarded as one of the Mechanical Powers, and so considering it we distinguish three kinds of lever, according to the position of the fulcrum with respect to the Power and Weight.

The first has the fulcrum between the power and the weight. In this case any amount of mechanical advantage may be gained, by making the arm upon which the power acts sufficiently long. A crow-bar used to lift great weights, a poker, a pair of scissors, are examples. Let us examine one of these; in the poker, the coals are the weight, the bar of the fire-place the fulcrum, the force exerted by the hand the power.

The second kind of lever has the weight between the fulcrum and the power. The oar of a boat is an example,

in which the water forms the fulcrum, the resistance of the boat applied at the rowlock the weight, and the power is applied by the hand of the rower.

The third kind has the point of application of the power between the fulcrum and the weight. The most interesting example is the human arm, when applied to lift a weight by turning about the elbow; here the fulcrum is the elbow, and the power is applied at the wrist by means of sinews, which exert a force when the muscles of the arm contract.

The mechanical conditions of each of these three classes may be thus expressed. If a represent the arm at which the power acts, b that at which the weight acts, then

$$Pa = Wb,$$

$$\text{or } \frac{P}{W} = \frac{b}{a}.$$

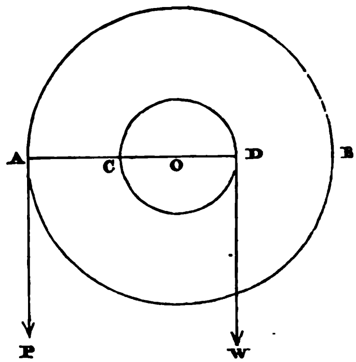
Hence mechanical advantage is gained or not, according as a is greater or less than b. It will be seen that in the first kind of lever advantage may or may not be gained, in the second it is always gained, in the third it is never gained. The human arm therefore acts at a mechanical disadvantage, but this is far more than compensated by the superior agility and neatness, which result from its actual construction.

(2) *The Wheel and Axle.*

21. This machine consists of two cylinders, having their axes coincident, the two cylinders forming one rigid piece, or being cut from the same piece; the larger is called the wheel, the smaller the axle. The cord by which the weight is suspended is fastened to the axle and coiled round it; the power acts, sometimes by a cord coiled round the wheel, sometimes by handspikes, as in the capstan, sometimes by handles, as in the windlass.

13

22. *To find the ratio of the Power to the Weight, when there is equilibrium upon the Wheel and Axle.*



Let AB, CD be the wheel and axle having the common centre O; P and W the power and weight, supposed to act by strings at the circumference of the wheel and axle respectively.

From the common centre O draw OA, OD to the points at which the cords supporting P and W touch the circumferences of the wheel and axle respectively; these lines will be perpendicular to the directions in which P and W act; hence, by the principle of the lever,

$$P \cdot AO = W \cdot OD,$$

$$\text{or} \quad \frac{P}{W} = \frac{OD}{AO} = \frac{\text{radius of axle}}{\text{radius of wheel}}.$$

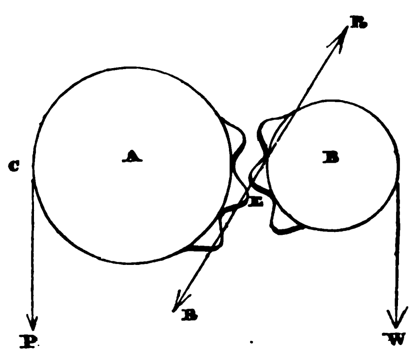
It is evident that, by increasing the radius of the wheel, any amount of mechanical advantage may be gained. It will also be seen that the principle of the wheel and axle is merely that of the lever, the peculiar advantage of the wheel and axle being this, that an endless series of levers (so to speak) are brought into play, which is essential to the practical use of the lever, when applied to such purposes as raising a bucket in a well, heaving an anchor, or the like.

(3) *The Toothed Wheel.*

23. One wheel may be made to act upon, or as it is called to *drive*, another by indenting the surface of each with teeth and fixing the centres at such a distance from each other that

the teeth come successively into contact. The proper form for the tooth of such wheels is a question of much complexity and care, which will not be entered upon here; we shall only investigate in general the relation of P to W, when there is equilibrium.

24. *To find the ratio of the Power to the Weight in the Toothed Wheel.*



Let A, B be the centres of the wheels, on the circumference of which the teeth are arranged, and suppose for simplicity's sake that P and W act at the circumferences of the wheels, and that the radii of the same are $r\,r'$ respectively.

Also let two of the teeth be in contact at E, and let R be the mutual pressure of the tooth in contact, which acts in the direction of the common normal to the surfaces of the teeth; and let $p\,p'$ be the lengths of the perpendiculars from A and B respectively on the common normal.

Then the wheel, of which the centre is A, may be supposed to be kept in equilibrium by the forces P and R tending to twist it in opposite ways; hence by the principle of the lever,

$$P \cdot r = R \cdot p.$$

Similarly, for the equilibrium of the other wheel,

$$W \cdot r' = R \cdot p'.$$

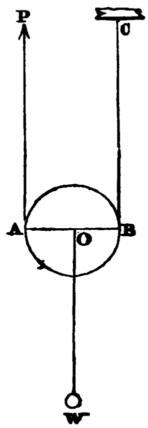
Hence $$\frac{Pr}{Wr'} = \frac{p}{p'},$$

or $$\frac{P}{W} = \frac{pr'}{p'r}.$$

(4) *The Pully.*

25. The Pully, in its simplest form, consists of a wheel, capable of turning about its axis, which may be either fixed or moveable. A cord passes over a portion of the circumference, and is supposed to be prevented by friction from sliding over its surface: if the axis of the pully is fixed, its only effect is to change the direction of the force exerted by the cord, but if it is moveable a mechanical advantage may be gained, as we shall see immediately. Combinations of pullies may be made in endless variety; we shall here consider only the simple moveable pully, and some of the more ordinary combinations.

26. *To find the ratio of the Power to the Weight in the simple moveable Pully.*



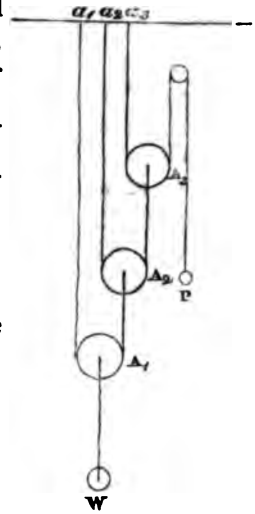
Let O be the centre of the pully, which is supported by a cord passing under it and attached to some fixed point C at one end, and stretched by the force P at the other. Suppose the weight W to be suspended from the centre O.

Join the points A, B, at which the contact of the cord with the pully commences, by a straight line AB, which will pass through the centre O. Then we may consider the mechanical conditions of the problem to be the same as those of a lever AB, kept in equilibrium about the fulcrum O by the force P at A and the tension of the string at B. But the tension of the string must be the same throughout, and is therefore equal to P. Hence the force at each end of the lever is P, and the resultant of these two parallel forces $2P$. But this resultant supports W;

$$\therefore\ 2P = W,$$

$$\text{or } \frac{P}{W} = \frac{1}{2}.$$

27. *To find the ratio of the Power to the Weight, in a system of Pullies in which each pully hangs by a separate string.* (First system of Pullies.)

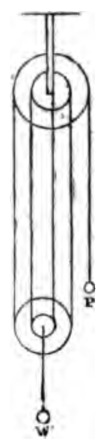


This system of pullies is represented in the figure. Suppose there are n pullies; then the tension of the string passing under the first $= \frac{W}{2}$ (by the property of the simple pully). The tension of the string passing under the second $= \frac{W}{2^2}$, and so on. That of the string under the last pully $= \frac{W}{2^n}$; but this must be equivalent to the power P.

$$\therefore \quad P = \frac{W}{2^n},$$

$$\text{or} \quad \frac{P}{W} = \frac{1}{2^n}.$$

28. *To find the ratio of the Power to the Weight, in a system of Pullies in which the same string passes round all the Pullies.* (Second system of Pullies.)

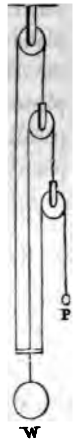


This system is represented in the figure. There are two blocks, the lower one moveable, and each containing a number of pullies. Since the same string goes round all the pullies, the tension throughout will be the same, and equal to the power P. Let n be the number of strings at the lower block, then the sum of their tensions will be nP, and we shall have,

$$nP = W,$$

$$\text{or} \quad \frac{P}{W} = \frac{1}{n}.$$

29. *To find the ratio of the Power to the Weight, in a system of Pullies in which all the strings are attached to the Weight.* (Third system of Pullies.)



The figure represents the system. The tension of the string by which P hangs is P; that of the next $= 2P$ (by the property of the simple pully); that of the next 2^2P, and so on. Let there be n strings, then the tension of the last $= 2^{n-1}P$, and the sum of all the tensions

$$= (1 + 2 + 2^2 + \dots\dots\dots\dots + 2^{n-1})\,P = W;$$

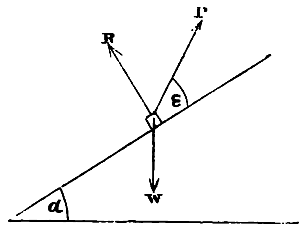
$$\text{or } \frac{P}{W} = \frac{1}{1 + 2 + 2^2 + \dots\dots\dots\dots + 2^{n-1}} = \frac{1}{2^n - 1}.$$

(5) *The Inclined Plane.*

30. By the inclined plane is meant a plane inclined to the horizon, and the problem is to find the force necessary to prevent a body placed upon it from sliding down under the action of its own weight. The plane is supposed smooth, and therefore, for reasons already explained (Art. 18, page 191), will exert a pressure on the body in the direction of a line perpendicular to its surface. We shall have to apply here the general principles of equilibrium before deduced, viz. that the sum of the forces acting on the weight, resolved in any two directions perpendicular to each other, must separately vanish: the two equations furnished by these conditions will enable us to determine not only the ratio of the power to the weight, but also the pressure of the weight on the plane.

31. *To find the ratio of the Power to the Weight, when there is equilibrium on the Inclined Plane.*

Let α be the inclination of the plane to the horizon; R the pressure of the plane on the weight, which will be perpendicular to the plane; and let ϵ be the angle which the direction of P makes with the plane. Then resolving the forces parallel and perpendicular to the plane, we have



$$P \cos \epsilon - W \sin \alpha = 0 \ldots\ldots\ldots (1)$$

$$R + P \sin \epsilon - W \cos \alpha = 0 \ldots\ldots\ldots (2).$$

Hence $$\frac{P}{W} = \frac{\sin \alpha}{\cos \epsilon}.$$

COR. 1. If the power acts parallel to the plane, $\epsilon = 0$, and

$$\frac{P}{W} = \sin \alpha.$$

COR. 2. If it is required to find the pressure R, we have, multiplying (1) by $\sin \epsilon$ and (2) $\cos \epsilon$, and subtracting,

$$R \cos \epsilon = W \cos (\alpha + \epsilon).$$

If $\epsilon = 0$, $$R = W \cos \alpha.$$

(6) *The Wedge.*

32. The wedge is a triangular prism, made of some hard substance, as steel, the edge of which is introduced between two obstacles, which it is our purpose to separate. When the edge is introduced, the wedge is driven forward by a violent blow, as from a hammer or the like, which generates an enormous force of momentary duration. We shall consider the wedge to be acted upon by a weight resting upon its head, but the principles of the investigation are

applicable to all cases, in whatever manner the pressure on the wedge is produced; we shall also suppose the wedge to be isosceles, and the obstacles on opposite sides of the wedge to be exactly similar. When the wedge is driven in between two obstacles, as for instance when applied to split the trunk of a tree, the obstacles have a tendency to fly together, owing, in the instance supposed, to the tenacity of the fibres, and this is the resistance which we consider, and which corresponds to the weight supported in the Mechanical Powers already treated of. In practice there will usually be a great amount of friction between the wedge and obstacles, but this, for the sake of simplifying the mathematical investigation, we shall omit.

33. *To find the ratio of the Power to the Resistance, when an isosceles Wedge is kept in equilibrium by the pressure of two obstacles symmetrically situated with respect to it.*

Let $2P$ be the force acting on the head of the wedge, a the semiangle of the wedge. Also let A be one of the points, at which the wedge is in contact with the obstacle; then there will be a mutual pressure at this point between the wedge and obstacle, which will be perpendicular to the side of the wedge, and which we will call R. There will be a similar pressure, on the other side of the wedge, from the other obstacle.

Fig. I.

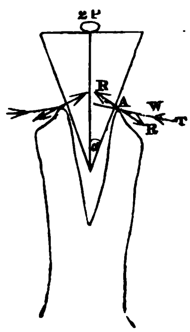
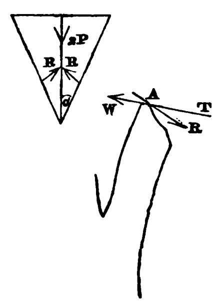


Fig. II.



Again, the point A of the obstacle is acted upon by the force which we call W, and which forms the resistance to motion. To determine the direction in which W acts,

we observe that, if A were made to move by the descent of the wedge, it would begin to move in some curve line, and that the tangent to that curve (AT) is the direction in which W acts. The position of this line AT is quite unknown, but if we denote the angle between it and the direction of R by i, we may be sure that i is in general small.

In Figure II. we have represented the wedge and obstacle separate from each other, in order to shew clearly the forces which respectively retain them in equilibrium. (The point A will of course require a third force to keep it in equilibrium, this will be perpendicular to AT and will arise from the pressure on the ground which supports the obstacle; it is omitted in the investigation, because its magnitude is a matter of no interest.)

Now the wedge is kept in equilibrium by the force $2P$ and the two forces R; hence, resolving in the direction of P, we have

$$2P = 2R \sin \alpha \ldots\ldots\ldots(1).$$

Again, the point A is kept in equilibrium, so far as tendency to motion in the direction AT is concerned, by W and the resolved part of R: hence

$$W = R \cos i \ldots\ldots\ldots(2).$$

From (1) and (2) we have

$$\frac{P}{W} = \frac{\sin \alpha}{\cos i}.$$

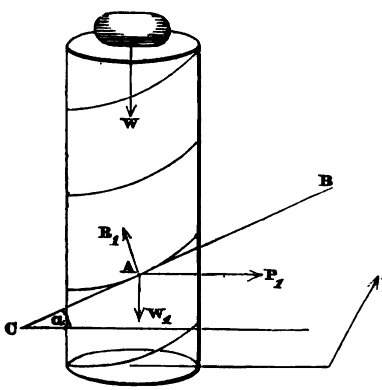
(7) *The Screw.*

34. The Screw may be conceived of as an inclined plane wrapped round a cylinder, or as a cylinder having on its surface a projecting thread inclined in all parts at the same given angle to the horizon. This cylinder fits into a block pierced

with an equal cylindrical aperture, on the inner surface of which is cut a groove the exact counterpart of the thread on the screw; hence we can cause the screw to enter the block only by making it revolve about its axis. Suppose the axis of the screw to be vertical, and a weight W to be placed upon it, then the screw would descend unless prevented from doing so by some other force; this force we suppose to be supplied by a power P acting in a horizontal direction, at the extremity of an arm of given length: this is nearly the mode in which the screw is actually applied to certain mechanical purposes, as to the bookbinder's press, and the like. In practice the friction between the thread of the screw and the block will generally be considerable, but for the sake of simplicity we shall consider every thing to be perfectly smooth.

35. *To find the ratio of the Power to the Weight in the Screw.*

Let the power P act at an arm a, and let r be the radius of the cylinder, α the inclination of the thread to the horizon.



Consider the equilibrium of any point A of the thread; suppose the portion of the thread on each side of A to be unwrapped, so as to assume the position of the straight line BC, inclined at an angle α to the horizon; then we may consider the point A as supported on a plane of inclination α, and acted upon by the pressure on the plane R_1, a horizontal force P_1, and a portion of the weight W, which we will call W_1; hence, resolving the forces along the plane, we must have

$$W_1 \sin\alpha = P_1 \cos\alpha;$$

in like manner, if we take another point of the thread, we must have

$$W_2 \sin\alpha = P_2 \cos\alpha,$$

and so on. Hence, taking into account all points of the thread, and adding together the equations, we shall have

$$(W_1 + W_2 + W_3 + \dots) \sin\alpha = (P_1 + P_2 + P_3 + \dots) \cos\alpha.$$

But $W_1 + W_2 + W_3 + \dots =$ the whole weight supported $= W$.

Also, $P_1 + P_2 + P_3 + \dots =$ the whole horizontal force supposed to act at the circumference of the cylinder, i.e. at an arm r.

But the horizontal pressure is caused by P acting at an arm a; hence, by the principle of the lever,

$$(P_1 + P_2 + P_3 + \dots)\, r = Pa;$$

$$\therefore\ W \sin\alpha = \frac{Pa}{r} \cos\alpha,$$

$$\text{or } \frac{P}{W} = \frac{r}{a} \tan\alpha.$$

We may put this result in a rather different form: thus,

$$\frac{P}{W} = \frac{2\pi r \tan\alpha}{2\pi a},$$

$$= \frac{\text{vertical distance between two threads}}{\text{circumference of circle described by } P}.$$

ON FRICTION.

36. When we attempt to make the surface of a body move upon that of another, with which it is kept in contact by pressure, there is in general a resistance to the motion, which is frequently sufficient to prevent motion altogether; the force of resistance is called *friction*.

Friction at any point of a surface always acts in the direction exactly contrary to that in which the point tends to move. Hence, when a particle is placed on a rough plane, the line in which the friction acts will lie in the plane. Also, if a body on an inclined plane, and under the action of some force, is on the point of *ascending*, the force of friction acts *downwards*; but if the weight of the body is so great that it is on the point of *descending*, the action of friction is *upwards*.

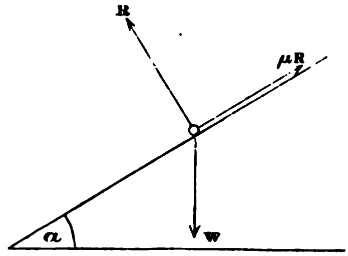
It has been shewn by experiment, that when a body is on the point of moving on the surface of another, and is only prevented from doing so by friction, then the force of friction bears to the normal pressure between the two bodies a ratio which depends only upon the constitution of the two bodies. In fact, let R be the normal pressure, then the friction will be expressed by μR, where μ is a quantity depending not on the pressure nor on the extent of the surfaces in contact, but only on the nature of the bodies; it has, for instance, a certain definite value for metal and wood, and so on. The quantity μ is called the *coefficient of friction*.

From what has been said it will appear, that in problems involving the pressure of one body on another, there will not be a greater number of unknown forces involved on the supposition of the bodies in contact being rough, than there would be on the hypothesis of their being smooth, provided we consider only the limiting circumstances of equilibrium, that is, when the body is on the point of sliding.

In consequence of the force of friction, systems of bodies in nature are not obliged to fulfil those exact conditions, which would be necessary if such a force did not exist. For example, a body would not retain its position on a smooth plane unless the plane were accurately horizontal, whereas a rough plane may be considerably inclined without disturbing the equilibrium of a body upon it.

37. We shall illustrate this by finding the angle at which a rough plane may be inclined, so that a body may just rest upon it without sliding.

Let a be the angle of inclination of the plane; W the weight of the body; R the normal pressure on the plane; μR the force of friction. Then resolving the forces perpendicular to the plane and parallel to it, we have these equations of equilibrium,



$$R = W \cos a,$$

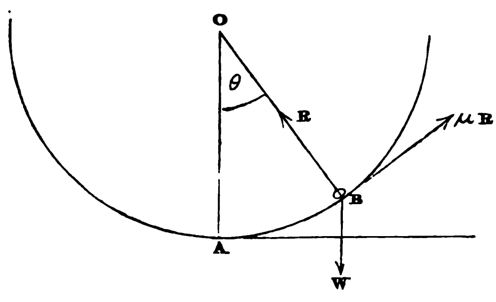
$$\mu R = W \sin a;$$

$$\therefore \quad \tan a = \mu.$$

This equation determines the limiting value of the inclination of the plane, for which equilibrium is possible; for any smaller value there will be equilibrium *a fortiori*.

38. Again, if the interior of a hemispherical bowl is smooth, a body cannot rest in it except at the lowest point; but if it be rough, there will be certain limits within which the equilibrium will be possible; let us determine those limits.

Let B be the highest position of the body possible, A the bottom of the bowl, O the centre; $AOB = \theta$. The direction



of the pressure R will pass through the centre, the friction μR

will be perpendicular to BO. Resolving the forces in the direction of BO and perpendicular to it, we have

$$R = W \cos\theta,$$

$$\mu R = W \sin\theta;$$

$$\therefore \quad \tan\theta = \mu.$$

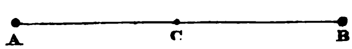
This equation determines the greatest possible angular distance of the body from the bottom of the bowl. The vertical height of the body above the lowest point A

$$= r(1 - \cos\theta) = r\left\{1 - \frac{1}{\sqrt{1+\mu^2}}\right\}.$$

Suppose, for instance, that $r = 1$ foot, and that $\mu = \frac{1}{4}$, which is its value for metallic surfaces, then the preceding expression becomes $1 - \frac{4}{\sqrt{17}} = .36$ of an inch, nearly.

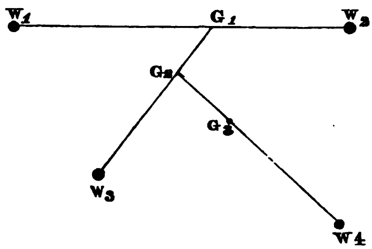
ON THE CENTRE OF GRAVITY.

39. If two equal heavy bodies A and B are connected by a fine rod, it is evident that the system will balance, if the middle point C of the rod be supported. And this will be the case in whatever position the system is placed, because the moment of A tending to twist the rod in one direction, will always be equal to the moment of B tending to twist it in the opposite. The point C, about which A and B will balance in any position, is called *the centre of gravity* of the bodies.



It may be shewn, that for every system of heavy particles there exists in like manner one point, and no more, such that, if it be fixed and the bodies rigidly connected with it, the system will rest in any position. This point is called *the centre of gravity of the system*.

40. *To shew that every system has a centre of gravity.*



Let W_1, W_2, W_3, W_4, ... be a system of particles, the weights of which are W_1, W_2, W_3, W_4, ... respectively: suppose W_1, W_2 joined by a rigid rod without weight, and divide the same rod in G_1 so that

$$W_1 G_1 : W_2 G_1 :: W_2 : W_1,$$

then W_1 and W_2 will balance in all positions about G_1, and if we suppose G_1 supported, the pressure upon the support will be $W_1 + W_2$.

Again, join $G_1 W_3$, and divide it in G_2 so that

$$G_1 G_2 : W_3 G_2 :: W_1 + W_2 : W_3;$$

then if we suppose the rod $W_1 W_2$ to rest upon the rod $G_1 W_3$, and G_2 to be supported, the pressure $W_1 + W_2$ at G_1 and W_3 at W_3 will balance about G_2. Hence the three bodies W_1, W_2, W_3, supposed rigidly connected, will balance in all positions about G_2.

Similarly we may find a point G_4, about which W_1, W_2, W_3, W_4 will balance in all positions, and so of any number of particles. Hence every system of particles has a centre of gravity.

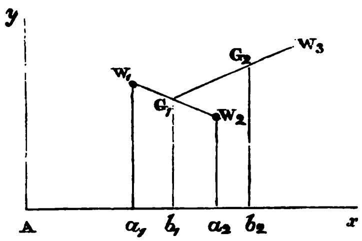
41. *A system can have only one Centre of Gravity.*

For suppose there are two, and let the system be so turned that the two centres of gravity lie in the same horizontal plane. Then the weights of the different particles of the system form a system of vertical forces, the resultant of which must pass through the centre of gravity, otherwise the system could not balance about that point; hence the said vertical resultant must pass through two points in the same

horizontal plane, which is absurd. Therefore there are not two centres of gravity.

42. It is manifest from the mode by which we proved the existence of a centre of gravity, that the tendency of a system of heavy particles to produce pressure, or to cause moment about any point, is the same as that of a single particle equal in weight to that of the whole system and situated at its centre of gravity. This is sometimes expressed by saying, that we may suppose a system *collected at its centre of gravity.*

43. *To find the centre of gravity of any number of particles in the same plane.*



Let W_1, W_2, W_3,...... be the weights of the particles; in the plane in which they lie, take any two straight lines Ax, Ay, at right angles to each other, and let h_1, h_2, h_3,.........be the distances of W_1, W_2, W_3, ... from the line Ax, and k_1, k_2, k_3, ... their distances from the line Ay; also let h, k the distances of the centre of gravity of the system from Ax, Ay respectively; then it is evident that if we find h and k we shall have solved the problem.

Join W_1W_2, and let G_1 be the centre of gravity of W_1, W_2; from W_1, W_2, G_1, draw W_1a_1, W_2a_2, and G_1b_1 perpendicular to Ax: then we have

$$W_1 \times W_1G_1 = W_2 \times W_2G_1;$$

but it is evident, from similar figures, that

$$W_1G_1 : a_1b_1 :: W_2G_2 : a_2b_1;$$

$$\therefore\ W_1 \times a_1b_1 = W_2 \times a_2b_1,$$

$$\text{or } W_1(Ab_1 - k_1) = W_2k_2;$$

$$\therefore Ab_1 = \frac{W_1k_1 + W_2k_2}{W_1 + W_2}.$$

If we consider another particle W_3, we may, in searching for the centre of gravity of the three W_1, W_2, W_3, suppose the two former to act together at their centre of gravity already found; hence, if G_2 be the centre of gravity of the three particles, and we draw G_2b_2 perpendicular to Ax, we shall have

$$Ab_2 = \frac{(W_1 + W_2)Ab_1 + W_3k_3}{W_1 + W_2 + W_3}$$

$$= \frac{W_1k_1 + W_2k_2 + W_3k_3}{W_1 + W_2 + W_3};$$

and so on for any number of particles. Hence we shall have

$$k = \frac{W_1k_1 + W_2k_2 \ldots\ldots\ldots\ldots + W_nk_n}{W_1 + W_2 \ldots\ldots\ldots + W_n};$$

and, in like manner,

$$h = \frac{W_1h_1 + W_2h_2 \ldots\ldots\ldots\ldots + W_nh_n}{W_1 + W_2 \ldots\ldots\ldots + W_n}.$$

44. *To find the centre of gravity of any number of particles not in the same plane.*

If we conceive three planes perpendicular to each other to be drawn, we shall solve the problem if we find the perpendicular distance of the centre of gravity from each of these planes. Let h_1, k_1, l_1, be the perpendicular distances of the particle W_1 from the three planes respectively, and so on of the other particles, and h, k, l, the perpendicular distances of the centre of gravity: then we may prove, in like manner as in the last proposition, that

$$h = \frac{W_1h_1 + W_2h_2 \ldots\ldots\ldots\ldots + W_nh_n}{W_1 + W_2 \ldots\ldots\ldots\ldots + W_n};$$

14

similarly,

$$k = \frac{W_1 k_1 + W_2 k_2 \dots\dots\dots\dots + W_n k_n}{W_1 + W_2 \dots\dots\dots\dots + W_n};$$

and so likewise,

$$l = \frac{W_1 l_1 + W_2 l_2 \dots\dots\dots\dots + W_n l_n}{W_1 + W_2 \dots\dots\dots\dots + W_n}.$$

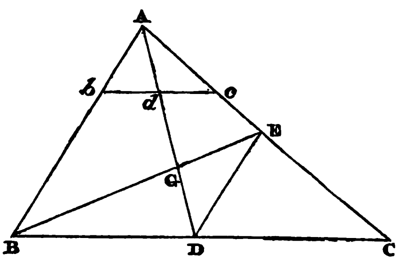
We shall now proceed to find the centre of gravity in some actual cases.

45. *To find the centre of gravity of a right line.*

The middle point will be the centre of gravity; for we may suppose the line to be divided into pairs of equal weights equidistant from the middle point, and the middle point will be the centre of gravity of each pair, and therefore of the whole system.

46. *To find the centre of gravity of a triangle.*

Let ABC be the triangle; bisect BC in D, and join AD; draw any line bdc parallel to BC; then it is evident that this line will be bisected by AD in d, and will therefore balance about d, in all positions; similarly all lines in the triangle parallel to BC will balance about points in AD, and therefore the centre of gravity must be somewhere in AD.



In like manner, if we bisect AC in E, and join BE, the centre of gravity must be in BE; hence G, the intersection of AD and BE, is the centre of gravity of the triangle.

Join DE, which will be parallel to AB. (Euclid, VI. 2.) Then the triangles ABG, DEG are similar.

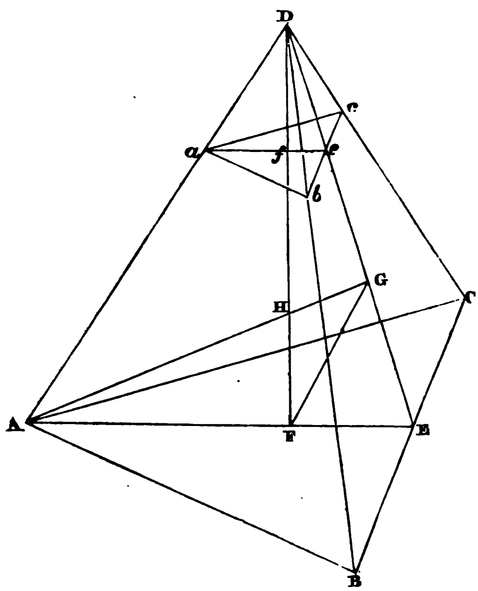
$$\therefore AG : GD :: AB : DE$$
$$:: BC : DC$$
$$:: 2 : 1,$$

or $AG = 2GD$,

and $\therefore AD = 3GD$.

Hence, if we join an angle of a triangle with the bisection of the opposite side, the point two thirds the distance down this line from the angular point is the centre of gravity of the triangle.

47. *To find the centre of gravity of a pyramid on a triangular base.*



Let $ABCD$ be the pyramid. Bisect BC in E; join AE; take $EF = \frac{1}{3} AE$, and join DF.

Suppose the pyramid to be made up of thin slices parallel to ABC, and let abc be one of them; let afe be the line, in which it is intersected by the plane DAE, e and f lying in bc and ae respectively. Then, by similar triangles,

$$be : eD :: BE : ED,$$

$$\text{also } ce : eD :: CE : ED;$$

$$\therefore \; be : ce :: BE : CE,$$

$$\text{but } BE = CE; \quad \therefore \; be = ce.$$

In like manner it may be shewn, that

$$fe : af :: FE : AF,$$

$$\text{but } AF = 2FE; \quad \therefore \; af = 2fe.$$

Hence f is the centre of gravity of the triangle abc. Similarly it will appear, that the centres of gravity of all slices of the pyramid made by planes parallel to ABC lie in DF, and therefore the centre of gravity of the pyramid is in that line.

Similarly, if we join DE, take $GE = \frac{1}{3} DE$, and join AG, the centre of gravity will be in AG; therefore H, the intersection of DF and AG, is the centre of gravity of the pyramid.

Now join GF, then by similar triangles,

$$\begin{aligned} HF : HD &:: GF : AD \\ &:: FE : AE \\ &:: 1 : 3; \end{aligned}$$

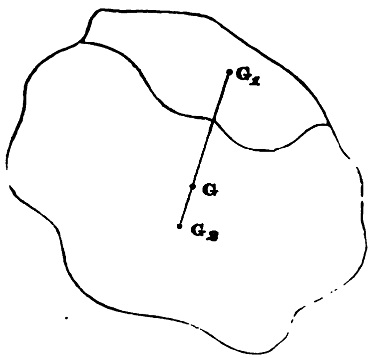
$$\therefore \; HF = \tfrac{1}{3} HD = \tfrac{1}{4} DF.$$

Hence, if we join the vertex of the pyramid with the centre of gravity of the base, and set off one fourth of this line from the latter point, we shall determine the centre of gravity of the pyramid.

Cor. 1. The same construction will hold for a pyramid upon a base of any form, since it may be divided into a number of pyramids on triangular bases.

Cor. 2. The centre of gravity of a solid cone will be found, by setting off one fourth of the axis measured from the centre of the base; for the base may be regarded as a polygon having an indefinite number of sides.

48. *Given the centre of gravity of a heavy body, and also that of a certain portion of it, to find the centre of gravity of the remainder.*



Let G be the centre of gravity of the body, W its weight: G_1 the centre of gravity of the given portion, W_1 its weight. Join $G_1 G$, and in that line produced take G_2, such that

$$G_2 G : G_1 G :: W_1 : W - W_1.$$

Then G_2 will be the centre of gravity required.

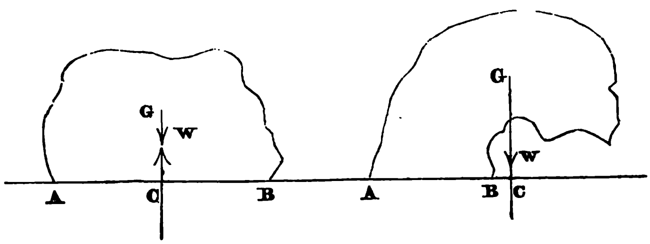
The preceding proposition is applicable to a variety of examples.

The following is one of the most important properties of the centre of gravity.

49. *When a body is placed upon a horizontal plane, it will stand or fall according as the vertical line through the centre of gravity falls within or without the base.*

Suppose the vertical line GC through the centre of gravity G, to fall within the base, as in fig. 1; then we may suppose the whole weight of the body to be a vertical pressure W acting in the line GC; this will be met by an equal and opposite

Fig. I. Fig. II.



pressure W from the plane on which the body is placed, and so equilibrium will be produced and the body will stand.

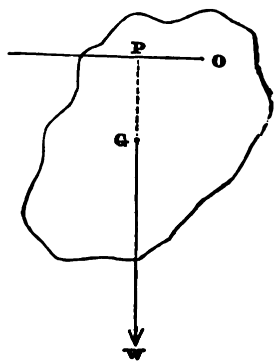
But suppose, as in fig. 2, that the line GC falls without the base; then there is no pressure equal and opposite to W at C, and therefore W will produce a moment about B, (the nearest point in the base to C,) which will make the body twist about that point and fall.

50. According to the proposition just proved, a body ought to rest without falling upon a single point, provided that it is so placed that the centre of gravity is in the vertical line passing through the point which forms the base. And in fact a body so situated would be, mathematically speaking, in a position of equilibrium, though practically the equilibrium will not subsist; this kind of equilibrium and that which is practically possible are distinguished by the names of *unstable*, and *stable*. Thus an egg will rest upon its side in a position of *stable* equilibrium, but the position of equilibrium corresponding to the vertical position of its axis is *unstable*. The distinction between *stable* and *unstable* equilibrium may be enunciated generally thus: Suppose a body or a system of particles to be in equilibrium under the action of any forces; let the system be arbitrarily displaced very slightly from the position of equilibrium, then if the forces be such that they tend to

bring the system back to its position of equilibrium the position is ***stable***, but if they tend to move the system still further from the position of equilibrium it is ***unstable***.

51. ***When a heavy body is suspended from a point about which it can turn freely, it will rest with its centre of gravity in the vertical line passing through the point of suspension.***

For let O be the point of suspension, G the centre of gravity, and suppose that G is not in the vertical line through O; draw OP perpendicular to the vertical through G, that is, to the direction in which the weight W of the body acts. Then the force W will produce a moment $W.OP$ about O as a fulcrum, and there being nothing to counteract the effect of this moment equilibrium cannot subsist.



Hence G must be in the vertical line through O, in which case the weight W produces only a pressure on the point O, which is supposed immoveable.

DYNAMICS.

1. We have now to treat of force, considered as producing *motion* in bodies. Our first business will be to explain accurately what we mean by the *velocity* of a body, and how it is measured.

2. The *velocity* of a body is the *rate* of its motion, or the degree of quickness with which it is moving: if of two bodies one passes in a given time over twice the distance passed over by the other, we say that the velocity of the first is twice as great as the velocity of the second.

Velocity may be *uniform* or *variable*. By saying that a body moves with uniform velocity, we mean that it moves through equal spaces in equal times; when the velocity is variable this is not the case.

Velocity when uniform is measured by the space passed over in a unit of time; when variable it is measured at any instant, by the space which would be passed over in a unit of time, if the body were to move during that unit with the velocity which it has at the proposed instant.

This requires some explanation. Let us first consider *uniform* velocity; in order to measure it, we first fix upon some unit of time, that is, some convenient period of time to which we may always refer, and by which we may measure other periods: the unit agreed upon is one *second*, so that in what follows, (unless the contrary be stated,) time will be measured by *seconds*; if any number, as 5 for example, should occur as representing time, it will be understood to mean $5''$ or 5 *seconds*. We may here also state that, in like manner, we find it convenient to agree upon a fixed unit of space, and that the unit we shall take will be one *foot*, so that (to take our former example) the number 5, if representing space, will mean 5 *feet*.

With these conventions our rule for measuring uniform velocity will be, that it is measured by the number of *feet* described in one *second*; and it does not require much consideration to perceive, that this is a proper mode of measuring velocity; for suppose two bodies are moving uniformly, and that one passes over 3 feet in a second and the other 5 feet, then the numbers 3 and 5 are proper representatives of the respective velocities of the bodies. Any other numbers in the same proportion would be equally proper expressions for the velocities, and the actual numbers of course depend upon the particular units we have chosen; thus, if in the case just supposed we had taken 2 seconds as the unit of time instead of 1 second, the values of the velocities would have been 6 and 10 instead of 3 and 5.

With regard to the mode of measuring *variable* velocity, it will be seen, that when the rate of a body's motion is changing from moment to moment, we cannot measure its velocity by the space which it passes over in a unit of time, because it will not have moved at the same rate during the whole of that unit of time. Hence we adopt the method already enunciated, and we measure the velocity of a body at any moment, not by any space actually described, but by the space which *would be* described in $1''$, if the body moved during that time with the velocity with which it is animated at the moment in question. In doing this, we are in fact only adopting a method which is in ordinary use; for when we speak of the velocity of a coach as 10 miles per hour, we do not mean to assert that the coach will pass over, or has passed over, 10 miles in any given hour, but only this, that *if* it were to proceed during an hour at the rate at which it was moving at the moment of our observation it *would* pass over 10 miles.

3. We shall in what follows usually denote time by the letter t, space by s, velocity by v. From what has been said, we shall be able to attach a distinct conception to each of these symbols; t will be the number of *seconds* in any time symbolized, s the number of *feet* in any space, and v the number of *feet* which are, or would be, described by a body in one *second*, according as the velocity is uniform or variable.

4. Prop. *If* s *be the space which a body, moving uniformly with a velocity* v, *describes in the time* t, *then* s = vt.

For v is the number of feet which the body passes over in 1″; and the body passes over equal spaces in equal times, therefore in t'' it describes vt feet, i.e. $s = vt$.

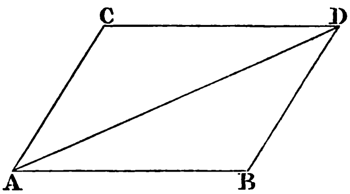
5. We may extend to velocity the convention respecting algebraical signs, which we have already found of use in the case of lines, angles, and statical forces. If we fix upon any point in a body's path, and consider the velocity of the body to be positive when its distance from that point is increasing, then we must regard the velocity as negative when that distance is diminishing: for instance, suppose a body is projected upwards from the earth's surface, and we regard the velocity during the ascent as positive, then when the body descends the velocity will be negative.

6. It is manifest that we may represent velocity, in the same way as we formerly represented force, by a straight line, the length of the line indicating the magnitude of the velocity, and the direction in which it is drawn that in which the body is moving; this we shall call representing a velocity in direction and magnitude.

7. When a body is animated simultaneously by two velocities having different directions, the body will evidently move in a direction intermediate to those two directions, in one of which it would have moved if animated by either velocity singly, and with a velocity which will in some manner depend on each of the two velocities, and may be called their ***resultant***. The actual problem of finding the resultant of two velocities we shall solve, by proving a proposition very nearly analogous to the parallelogram of forces, and which we shall call the *parallelogram of velocities*.

8. Prop. *If two velocities, with which a particle is simultaneously animated, be represented in direction and magnitude by two straight lines drawn from the particle,*

the resultant velocity of the particle will be represented in direction and magnitude by the diagonal of the parallelogram described upon those two straight lines.



Let A be the particle, AB AC the lines representing the two component velocities, AD the diagonal of the parallelogram $ABDC$ described upon them.

Then, under the influence of the velocity AB only, the particle would at the end of $1''$ be at B, and under the influence of AC it would be at C; suppose now, that instead of moving for $1''$ under the influence of the two velocities, the particle moves for $1''$ under the influence of AB, and then for $1''$ under the influence of AC, its place at the end of the time will be the same as upon the former hypothesis. But on this supposition the particle will be at the end of the first second at B, and (drawing BD parallel and equal to AC) it will be at the end of the next at D; hence, under the influence of AB and AC jointly, the particle will be at the end of $1''$ at D, that is to say, AD represents the resultant velocity in direction and magnitude.

Obs. In this proof we have, for simplicity's sake, spoken of the velocities as if they were uniform; the same proof however applies, with a change of phraseology, to variable velocities.

9. It follows from the proposition just proved, that, if a particle be moving with a velocity v in a direction making an angle θ with a given line, we may conceive the particle to be animated by two velocities, $v \cos\theta$ in the direction parallel to the given line, and $v \sin\theta$ in the direction perpendicular to it. This is called *resolving* a velocity.

And hence we may compound any number of velocities and find their resultant, by a process exactly similar to that made use of for forces, and given at length in Art. 14, page 185.

Having said thus much respecting velocity, we shall now proceed to treat of *force*, considered dynamically.

10. We shall begin by enunciating the following general property of matter, which is known as the *First Law of Motion*.

A body under the action of no external force will remain at rest, or move uniformly in a straight line.

With regard to the *meaning* of this law, it is intended to assert that there is in matter no tendency to motion of one kind more than another, or indeed of any kind; that matter is purely inert, and the cause of any motion which a body may have, is to be sought not in the properties of the body itself, but in external influences. The property which matter possesses, of moving only with the velocity it has ~~re~~acquired from the action of external force, is called its *inertia*; hence we may say briefly, that the first law of motion asserts the *inertia* of matter.

11. As to the *proof* of this law, we may obtain some hint of its truth by observing, that the more nearly we make the circumstances of a body agree with those supposed, the more nearly is the law verified: for instance, according to the law the destruction of the velocity of a body moving along a dead level ought to be wholly due to friction, and we do in fact find that the more we guard against friction the longer the body will continue in motion, and on a railway, where the friction is very much diminished, the distance to which a train will proceed after the steam has been turned off, is very great indeed. Other examples will suggest themselves to the student, or perhaps he may think the law so simple as not to require illustration; he must remember however that the want of a clear perception of its truth was for a long time a bar to progress in dynamical science, because men, misled by terrestrial phenomena, considered it necessary to inquire what force was necessary to keep a body in a certain uniform state of motion. A satisfactory proof of the truth of this, as well as of other laws which we shall meet with hereafter, arises from the fact

of the accurate agreement with fact of calculations, many and complicated, which are based upon it. Perhaps, however, the mind, which has dwelt long on the subject, will see the truth of the law as necessarily involved in the idea of matter, and as having therefore an axiomatic character more convincing than any proof founded upon the agreement of calculation and experiment: we must not however pursue a remark, which would lead us into discussion unsuited to the character of the present elementary work.

12. Dismissing then the question of the nature of the proof of the first law of motion, and supposing its truth to be established, we learn from it that whenever a body is moving with a velocity varying either in direction or in magnitude, we are to conclude that the body is acted upon by some extraneous force. Confining ourselves, for distinctness of conception, to the case of a body moving in a straight line but not uniformly, we are to conclude that the change of velocity from one moment to another is due to the action of some force upon the body; hence the rate of this change must in some way be an index of the intensity of the force: but here the question occurs,—Are we to estimate the force acting on the body solely by the change of velocity, or are we to take into account also the quantity of matter contained in the body, for it is an obvious fact that a given force (as a blow for instance) will generate a greater velocity in a body as its weight is less? The answer is, that we may estimate the force either way, that is, we may estimate it either by the velocity generated only, or we may take into account not only the velocity generated, but also the quantity of matter moved; in the former case we call the force the *accelerating* force acting on the body, in the latter the *moving* force. As the distinction between these two ways of estimating force is of first importance to a clear understanding of dynamical principles, we shall enunciate it as plainly as we can.

13. *Force considered with reference to velocity generated only, and not to the quantity of matter moved, is called* accelerating *force.*

Force considered with reference to the quantity of matter moved, as well as the velocity generated, is called moving *force.*

14. The exact relation between accelerating and moving force will be for our consideration presently: the former will just now occupy our attention, and we must consider carefully the mode of measuring it. We will observe, by the way, that the term *accelerating* is not used as opposed to, but rather as inclusive of, *retarding* force, the latter being supposed to differ from the former merely in its algebraical sign; and hence, when we speak of a force *generating* a velocity in a body, we intend the term to include forces which *destroy* velocity.

Accelerating force may be *uniform* or *variable;* it is called *uniform* when equal velocities are generated in equal times, *variable* when this is not the case.

Accelerating force is measured, if uniform, by the velocity generated in a unit of time; if variable, by the velocity which would be generated in a unit of time if the force remained constant during that unit.

Let us illustrate the mode of measuring accelerating force, by reference to the case of the earth's attraction: we observe that bodies fall with a variable velocity, hence we know that they are acted upon by some force; this force we are led to attribute to an attraction residing in the earth, and which we call *gravity;* again, we observe, (or at least we will suppose it to be observed,) that the increments of velocity of a falling body in equal times are equal, hence we conclude that gravity is a *uniform* force. Let us consider how it is to be measured; a body in 1″ is found to fall from rest through 16.1 feet, at the end of 1″ it is found to be moving with such a velocity that it would, if it continued to move with the velocity which then animates it, pass over in the next second 32.2 feet. Hence 32.2 feet is the measure of the velocity which has been generated in 1″, and is therefore the measure of the accelerating force of gravity. It may

be mentioned here, that this quantity is usually denoted by the letter g.

It appears then that 32.2 feet is the measure of the earth's attraction, and in making use of this result the student is requested to bear carefully in mind all the conventions upon which it depends. We have assumed that the accelerating force of the earth's attraction on all bodies is the same; an experimental proof of this is supplied by the fact, that under the exhausted receiver of an air pump all bodies fall equally rapidly; the difference of velocity of falling bodies in air is due entirely to the different action of the air upon them.

15. PROP. *If* v *be the velocity generated by a uniform accelerating force* f *in the time* t, *then will* v=ft.

For f is the expression for the velocity generated in 1″, and since the force is uniform an equal velocity is generated in each second, therefore the velocity generated in t'' is ft; i.e. $v = ft$.

Thus in the case of the earth's attraction, the velocity generated in 1″ in a falling body is 32.2 feet, therefore the velocity generated in 2″ is 64.4 feet; by which we mean, that if a heavy body be let fall, it will at the end of 2″ be moving with such a velocity that, were that velocity continued constant through 1″, the body would move through 64.4 feet.

16. We have said that force is measured by the velocity generated in a unit of time, but now we must observe further that there is a certain kind of force which cannot be so estimated. This kind of force is what we term *impulsive*, under which name we include all forces which are of the nature of a sudden blow, applying the name of *finite* to all other forces, such as that arising from the gravitation of bodies. We will first give formal definitions of these two classes of force, and then attempt to explain clearly the difference between them, and the necessity for different modes of measuring their effects.

Def. A *finite* force is one which requires a finite time to generate a finite velocity.

Def. An *impulsive* force is one which generates a finite velocity in an indefinitely short time.

17. Let us consider what takes place, when an impulsive force results from the impact of two bodies upon each other. For distinctness of conception, let us suppose that the impact is that of two ivory balls: when the balls strike each other, they appear to fly asunder instantaneously, but in reality a rather complicated action takes place between them; the first effect after impact is a compression of each of the balls, very slight of course, but nevertheless necessarily existing unless the balls be absolutely rigid, which no substance in nature is; after the compression has ceased, a restitution of the figure of the balls commences in consequence of their *elasticity*, and when the form of the balls is restored they separate, and their action on each other ceases. The whole action first described necessarily occupies a certain space of time and cannot be instantaneous, and there is nothing in the nature of the forces considered to make them differ essentially from those which we have denominated *finite;* nevertheless, if we attempt to measure them in the same way we are met by this insuperable difficulty, that we know nothing of the laws according to which the action we have described takes place, we cannot observe the action of the forces, because for all purposes of observation that action is instantaneous; hence we distinguish these *impulsive* forces as a new class of forces, not because they are physically different from those which we call finite, but because, since they generate a finite velocity in a time so short as to be considered indefinitely small, we are compelled to measure them in a different way; and we measure the accelerating force of an impulse, not by the velocity which would be generated in a unit of time, but by the whole velocity actually generated.

The measure of the *moving force* of an impulse will be considered, when we come to speak in general on the connection between accelerating and moving force.

18. The first Law of Motion would enable us, with the help of such mathematical calculation as we shall hereafter employ, to solve all problems of rectilinear motion; but we have at present no means of determining the motion of a body which is moving under the action of various forces in different directions, or of a body which being in motion is acted upon by a force not in the direction of the motion. That which we require is furnished us by the *Second Law of Motion*, which we shall at once enunciate.

19. *When any number of forces act upon a body in motion, each produces its whole effect in altering the magnitude and direction of the body's velocity, as if it acted singly on the body at rest.*

The application of this law is as follows: At any moment a body has a certain velocity, also if it were at rest each of the forces would, acting separately, generate in it a certain velocity in a certain direction; we must suppose that the body has all these velocities simultaneously, and if they be compounded by the rule of the parallelogram of velocities, the resultant velocity will be the actual velocity of the body.

20. With regard to the proof of the second Law of Motion, we may appeal to such experiments as the following: A ball let fall from the mast of a ship in motion, will fall at the foot of the mast, notwithstanding the onward course of the ship: in this instance, the velocity of the ball is compounded of that which it has in consequence of the motion of the ship, and of that impressed upon it by the action of gravity; for before the ball is let fall, it has the same horizontal velocity as the ship, and when let fall, it does not lose this velocity, but only has it compounded with another, viz. that due to gravity, and as the ship also retains the same horizontal velocity, (for it is supposed to be sailing uniformly,) the ball falls, with reference to the deck, exactly as though the vessel were at rest.

Illustrative of this law also are such facts as these; that we walk with the same ease in different directions along the

earth's surface, although we know the earth to be not at rest; that we can walk across a railway carriage in rapid motion; and doubtless many others will suggest themselves to the mind of the thoughtful student. Perhaps the most satisfactory is this; the time of oscillation of a pendulum is independent of the plane in which it vibrates, notwithstanding the earth's motion.

Further experimental proof arises from the fact of calculations, based upon this law, leading in the most delicate cases to correct results. Perhaps however it may be said, that when the idea of force has been completely seized and thoughtfully considered, the truth of the second Law of Motion will present itself to the mind in a form rather axiomatical than experimental; but this is a question on which we shall not here raise a discussion.

21. Hitherto we have considered force only as measured by the velocity which it generates, in other words, we have been concerned with *accelerating force:* we must now examine the proper mode of estimating force, when we take into account the quantity of matter moved, or when we regard it as *moving force.*

The quantity of matter in a body is usually termed its *mass;* but how are we to define the term mass? If we were to conceive matter to be made up of ultimate atoms, which atoms should be all precisely alike in magnitude and in all their qualities, then we might measure the mass of a body by the number of atoms it contains; but this is not a practicable method, and we are compelled to measure the mass of a body by its effects; we measure the mass of a body by its *weight*, or we consider two bodies to be of the same *mass* if they are of the same *weight.* In this mode of estimating mass, it is assumed that the attraction of the earth on all particles is the same, that is, that the attraction is not dependent upon the nature of the matter attracted, an assumption which appears to be justified by the fact already alluded to, that bodies of all kinds fall with equal

velocity under the exhausted receiver of an air-pump, and of the truth of which Newton assured himself by a variety of experiments. We can only define the mass of a body then, as being a quantity proportional to its weight, so that, if W be the weight of a body and M its mass,

$$W \propto M,$$

$$\text{or } W = C \,.\, M,$$

where C is some constant quantity, the numerical value of which will depend upon circumstances to be presently explained.

22. Before proceeding further, we must define the term *Momentum*, which we shall use frequently.

DEF. The *momentum* of a particle of matter is its mass multiplied by its velocity; the momentum of a body, or system of particles, is the sum of the products of the masses of the particles by their respective velocities.

The phrase *quantity of motion* is sometimes used instead of *momentum*.

23. Let us now inquire concerning the method of measuring *moving force*; in other words, when pressure communicates motion to a body, what will be the proper measure of its effect? The answer is given by the *third Law of Motion*, which we may enunciate as follows:

When pressure produces motion in a body, the momentum generated in a unit of time, supposing the pressure constant, or which would be generated supposing the pressure variable, is proportional to the pressure.

24. The result of this law is, that as *velocity* generated is the measure of *accelerating* force, so *momentum* generated is the measure of *moving* force. And if P be a uniform pressure, v the velocity generated in the time t, in a body the mass of which is M, we shall have

$$Mv = Pt.$$

Suppose f to be the accelerating force corresponding to the pressure P, then we have

$$v = ft,$$

hence, comparing the two last formulæ,

$$P = Mf,$$

or, moving force = mass × accelerating force.

Thus the third Law of Motion teaches us how, when a pressure is given, to deduce the accelerating force due to it, for we have only to divide the pressure by the mass moved, and we have the accelerating force required. Let us take the example of the pressure caused by the weight of a body; let W be the weight, M the mass, then we know that the accelerating force is that which was before denoted by g, hence by the third Law of Motion we must have

$$W = Mg,$$

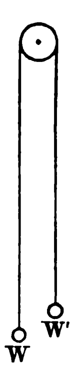
and therefore the constant C in the formula of Art. 21. is to be replaced by g, the accelerating force of the earth's attraction.

25. This Law of course includes *impulsive* forces, and we may therefore say that an impulsive force is measured by the whole momentum generated by it. (See Art. 17.)

26. We may exhibit the third Law of Motion from a slightly different point of view, by saying, that it establishes a relation between the statical measure of force, which is the pressure produced, and the dynamical measure, which is the velocity generated.

27. With regard to the truth of the third Law of Motion, it may be considered either to rest on experiment, confirmed by the coincidence with fact of innumerable calculations founded upon it, or perhaps to be deducible without experiment from previously established principles.

We shall content ourselves with giving an account of one experiment, divested of all the refinements by means of which *Atwood* was able, with a machine constructed for the purpose, to obtain great accuracy of result.



Let W and W' be two weights, of which W is the greater, connected by a fine string passing over a small fixed pully. Then W will descend, and as W and W' are pulling in opposite directions, the weight producing motion is $W - W'$, and the weight moved is $W + W'$. The advantage of this experiment is, that we can alter the difference between the weights ($W - W'$) as much as we please, while the weight moved ($W + W'$) remains the same. Now by numerous experiments it was determined, that the ratio $\frac{W - W'}{W + W'}$ is in all cases proportional to the accelerating force, or that the pressure producing motion is proportional to the product of the weight moved and the accelerating force, or (which is the same thing) the product of the mass and the accelerating force: which result is in accordance with the third Law of Motion.

28. We are now prepared with Dynamical principles, which we shall proceed to apply to the cases, of a particle under the action of uniform forces, and of the collision of bodies. A more general application of the principles would require the use of a higher calculus than any with which the student is supposed to be acquainted.

ON THE RECTILINEAR MOTION OF A SINGLE BODY CONSIDERED AS A PARTICLE, UNDER THE ACTION OF A UNIFORM FORCE.

The following is the fundamental proposition of this branch of Dynamics.

29. Prop. *If* s *be the space described from rest, in the time* t, *by a body under the action of a uniform accelerating force* f, *then* $s = \frac{ft^2}{2}$.

Suppose the time to be divided into n equal intervals, each equal to τ, so that $t = n\tau$; then the velocity of the body at the end of the 1st, 2nd, 3rd.... intervals will be $f\tau$, $2f\tau$, $3f\tau$... respectively; (Art. 15.) Now suppose the body, instead of moving as it actually does, to move through the whole of each interval of time with the velocity which it had at the *beginning* of the interval, and let s_1 be the space which would be described in the time t on this hypothesis; again, suppose the body to move through each interval with the velocity which it has at the *end* of the interval, and let s_2 be the space described on this hypothesis; then

$$s_1 = 0.\tau + f\tau.\tau + 2f\tau.\tau + \dots + (n-1)f\tau.\tau$$

$$= f\tau^2 \frac{n(n-1)}{2} = \frac{ft^2}{2}\left(1 - \frac{1}{n}\right),$$

$$s_2 = f\tau.\tau + 2f\tau.\tau + 3f\tau.\tau + \dots + nf\tau.\tau$$

$$= f\tau^2 \frac{n(n+1)}{2} = \frac{ft^2}{2}\left(1 + \frac{1}{n}\right).$$

Now it is manifest that s is intermediate in value to s_1 and s_2, but when we suppose the interval τ to be indefinitely small, and n indefinitely great, we bring the two hypothetical cases which we have supposed to coincide with the real motion of the body, and in this case $\frac{1}{n}$ is indefinitely small, and

$$s_1 = s_2 = \frac{ft^2}{2};$$

$$\therefore \text{ also, } \quad s = \frac{ft^2}{2}.$$

30. We shall give another proof of the same proposition, which, though in some respects inferior to that which precedes, is worthy of the student's attention on account of its brevity.

Let A be the point from which the body starts, B its place at the end of the time t, so that $AB = s$: then the velocity of the

body at $B = ft$. Now suppose the body to start from B with a velocity ft, and to proceed towards A, the force f acting in the same direction as before, and therefore *retarding* its motion: at the end of the time t the body would, if no force acted, have described a space ft^2; also we know that the space, through which the force f would carry it, is s; hence the space through which the body will move in the time t, when it starts with a velocity ft, and is retarded by a force f, must be $ft^2 - s$. But at the end of the time t the body will be at A, since the force f will destroy the velocity ft, in exactly the same space which it required to generate it. Hence we have

$$ft^2 - s = BA = s;$$

$$\therefore \quad s = \frac{ft^2}{2}.$$

COR. If v be the velocity of the body at the end of the time t, $v = ft$; $\therefore s = \dfrac{v^2}{2f}$, or $v^2 = 2fs$. Hence, when a body falls under the action of gravity, this being a uniform force, the velocity $\propto$ the time, and the square of the velocity $\propto$ as the space described.

31. If the body, instead of starting from rest, begins to move with a given velocity, the formulæ of the preceding article are easily adapted.

For let V be the initial velocity, v the velocity at the time t, s the space described. Then the velocity at the time t is the original velocity $\pm$ that which is generated or destroyed by the force, or

$$v = V \pm ft,$$

the upper or lower sign being used, according as the force accelerates or retards.

And the space will be that, which would have been described by the body moving uniformly with a velocity $V \pm$ that

which it would describe under the action of the force only, or

$$s = Vt \pm \frac{ft^2}{2},$$

the upper or lower sign being taken according to the same rule as before.

Cor. We can obtain v in terms of s: for

$$\begin{aligned} v^2 &= (V \pm ft)^2 \\ &= V^2 \pm 2f\left(Vt \pm \frac{ft^2}{2}\right) \\ &= V^2 \pm 2fs. \end{aligned}$$

32. For distinctness' sake, we will here collect in a tabular form the formulæ which have been proved, and which may be applied to a variety of examples.

When the body starts from rest,

$$\left.\begin{aligned} s &= \frac{ft^2}{2} \\ v &= ft \\ v^2 &= 2fs \end{aligned}\right\} \text{......}(A).$$

When the body starts with an initial velocity V,

$$\left.\begin{aligned} s &= Vt \pm \frac{ft^2}{2} \\ v &= V \pm ft \\ v^2 &= V^2 \pm 2fs \end{aligned}\right\} \text{......}(B),$$

the upper or lower sign to be taken, according as f accelerates or retards the motion.

33. These formulæ possess peculiar interest, because they apply to the case of a body falling under the action of gravity. We have before said, that gravity is a uniform force,

at least it is sensibly so for small distances above the earth's surface; we may here remark in addition, that in consequence of the earth not being accurately spherical, the force of gravity is not exactly the same at all points of the earth's surface, but varies slightly with the latitude; we may consider, however, that for all ordinary purposes its magnitude is measured by the quantity $g = 32.2$ feet. The value of g is not determined by direct experiment, but may be most accurately found by observations of the pendulum, according to principles hereafter to be developed.

34. We shall now illustrate the formulæ (A) and (B) by some examples of falling bodies.

Ex. 1. Two bodies are let fall at an interval of $1''$, to find how far they will be apart after the lapse of $4''$ from the fall of the first.

Let $s\,s'$ be the spaces through which they have respectively fallen, then

$$s = \frac{g}{2}4^2 = \frac{g}{2} \times 16,$$

$$s' = \frac{g}{2}3^2 = \frac{g}{2} \times 9;$$

$$\therefore\ s - s' = \frac{g}{2} \times 7 = 16.1 \times 7 = 112.7 \text{ feet, the distance required.}$$

Ex. 2. When two bodies are let fall from the same point, but not at the same moment, the distances between them at the end of successive equal intervals increase in arithmetic progression.

Let $s\,s'$ be distances, described by the two bodies respectively in the time t from the fall of the first, τ the interval of time between the starting of the two, then

$$s = \frac{gt^2}{2},$$

$$s' = \frac{g}{2}\{t - \tau\}^2,$$

$$\therefore\ s - s' = \frac{g}{2}\{2t\tau - \tau^2\};$$

$s - s'$ is the distance between the bodies at the time t, c
this d, and let $d_1\ d_2$...be the distances corresponding to t
times $t + t_1$, $t + 2t_1$

$$\therefore\ d = \frac{g}{2}\{2t\tau - \tau^2\},$$

$$d_1 = \frac{g}{2}\{2(t + t_1)\tau - \tau^2\},$$

$$d_2 = \frac{g}{2}\{2(t + 2t_1)\tau - \tau^2\},$$

$$\&c. = \&c.;$$

$$\therefore\ d_1 - d = gt_1\tau = d_2 - d_1 = \&c.;$$

that is, the quantities dd_1d_2 ... &c., are in arithmetic p
gression.

Ex. 3. A stone is thrown into a well, and it is obser
that 2″ elapses before the sound of its striking the bottom
heard; neglecting the time occupied by the transmission
the sound, find the depth of the well.

Let s be the depth;

$$\text{then } s = \frac{g}{2}2^2 = 2g = 64.4 \text{ feet.}$$

Ex. 4. A ball is projected upwards with a given veloci
it falls, and on striking the earth rebounds with a loss
$\left(\frac{1}{n}\right)^{\text{th}}$ part of its velocity; find to what height it will r
after any given number of rebounds, and the whole spa
which will be described by the ball before coming to rest.

Let V be the velocity, then the height to which it rises $= \dfrac{V^2}{2g}$.

When it reaches the earth its velocity is V, therefore the velocity of rebound is $V\left(1 - \dfrac{1}{n}\right)$ by hypothesis, and the height to which it rises $= \dfrac{V^2}{2g}\left(1 - \dfrac{1}{n}\right)^2$.

In like manner, the height to which it rises after the 2nd rebound $= \dfrac{V^2}{2g}\left(1 - \dfrac{1}{n}\right)^4$.

And similarly, it will be seen, that the height after the p^{th} rebound $= \dfrac{V^2}{2g}\left(1 - \dfrac{1}{n}\right)^{2p}$, which is the answer to the first part of the question.

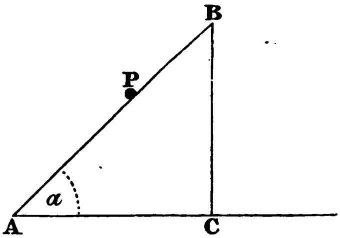
The whole space described by the ball

$$= \frac{V^2}{g}\left\{1 + \left(1 - \frac{1}{n}\right)^2 + \left(1 - \frac{1}{n}\right)^4 + \dots \textit{ad infinitum}\right\}$$

$$= \frac{V^2}{g}\frac{1}{1 - \left(1 - \dfrac{1}{n}\right)^2} = \frac{V^2}{g}\frac{n^2}{n^2 - (n-1)^2} = \frac{V^2}{g}\frac{n^2}{2n - 1}.$$

35. The formulæ (A) and (B) are applicable, with a slight modification, to the case of a body falling down a smooth inclined plane. In this case we must conceive the force of gravity to be resolved into two parts, one parallel to the plane, the other perpendicular to it; the latter will produce pressure on the plane, the former will accelerate the motion of the body and is the only part with which we shall be here concerned. Let a be the inclination of the plane, then the resolved part of g parallel to the plane will be $g \sin a$, and this quantity must be used for f in our fundamental formulæ.

Ex. 1. To find the velocity acquired by a body in falling down a given inclined plane.



Let AB be the inclined plane, α its inclination, P the place of the body at a given time, $BP = s$, v = the velocity at P; then we have by our formula,

$$v^2 = 2g \sin \alpha \times s,$$

which gives the velocity. If we draw BC perpendicular to the horizontal line through A, and call V the velocity at A, we have

$$V^2 = 2g \,.\, AB \sin \alpha$$
$$= 2g \,.\, BC.$$

Hence the velocity is the same at A, as if the body had fallen through the vertical space BC: that is to say, the velocity generated by gravity, depends upon the vertical space through which it is allowed to act, a result which might perhaps have been anticipated, and which was in fact assumed by Galileo.

Ex. 2. A body is projected upwards, along an inclined plane, with a given velocity, find how high it will ascend, and the time of ascent.

If v be the velocity at the time t, when it has ascended through a space s, α the angle of the plane, and V the given velocity of projection, we have

$$v^2 = V^2 - 2gs \sin \alpha\,;$$

when $v = 0$ the body will stop, and the distance required will be given by the equation,

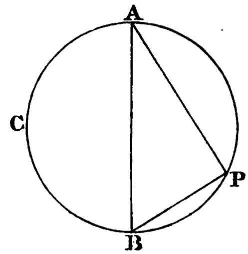
$$s = \frac{V^2}{2g \sin \alpha}.$$

For the time, we have,

$$v = V - gt \sin \alpha,$$

and the body stops when

$$t = \frac{V}{g \sin \alpha}.$$



Ex. 3. Let ACB be a circle in a vertical plane, A its highest point; the time of descent down all chords, considered as inclined planes, will be the same.

Let AP be any chord, α its inclination to the horizon; draw the vertical diameter AB, and join BP, then $ABP = 90^0 - BAP = \alpha$: now if t be the time of descent down AP, we shall have

$$AP = g \sin \alpha \frac{t^2}{2};$$

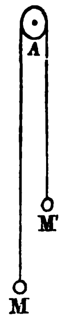
$$\text{but, } AP = AB \sin \alpha,$$

$$\therefore AB = \frac{gt^2}{2},$$

and $t = \sqrt{\frac{2AB}{g}}$, which is independent of α, and is therefore the same for all chords.

ON THE MOTION OF TWO FALLING BODIES CONNECTED BY A STRING.

36. We have distinguished problems of this class from those which we have been recently engaged in considering, because in those we were able to regard gravity only as an accelerating force, whereas in these we must consider the mass moved, and deduce the accelerating force by the third Law of Motion.



Ex. 1. Suppose we have two bodies, the masses of which are M and M' respectively, connected by a fine string passing over a smooth peg or small pully A, the only effect of which is to change the direction of the string.

Let M be greater than M'; then the moving force producing motion is the difference of the weights, or $Mg - M'g$, and the whole mass moved is $M + M'$, consequently the accelerating force is $\frac{M - M'}{M + M'} g$.

Hence, if x be the distance of M from A at the time t, a the distance when the motion commenced, we shall have

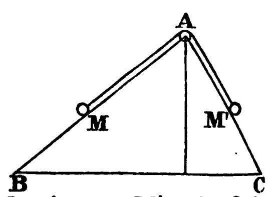
$$x = a + \frac{M - M'}{M + M'} \frac{gt^2}{2}.$$

Ex. 2. Let us take a numerical illustration of the last example.

Suppose $M = 2M'$, and $a = 0$, to find how far M will descend in $1''$. We have,

$$x = \frac{2 - 1}{2 + 1} \frac{g}{2} = \frac{16.1}{3} = 5.3 \text{ feet}.$$

Ex. 3. Two weights are placed upon two opposite inclined planes, and connected by a fine string which passes over a small pully at the highest point of the planes; to determine the motion.



Let AB, AC be the two planes, α, β their respective inclinations.

Then the part of the weight Mg which is effective in producing motion is $Mg \sin \alpha$, and that of $M'g$ is $M'g \sin \beta$; the difference of them or $Mg \sin \alpha - M'g \sin \beta$ is the moving force; the mass moved is $M + M'$; therefore the accelerating force is $\frac{M \sin \alpha - M' \sin \beta}{M + M'} g$.

If x be the distance of M from A at the time t, a the distance at the beginning of the motion,

$$x = a + \frac{M \sin \alpha - M' \sin \beta}{M + M'} \frac{gt^2}{2}.$$

Ex. 4. A numerical illustration of the preceding.

Suppose $M = 3M'$, $\alpha = 30^0$, $\beta = 60^0$: then,

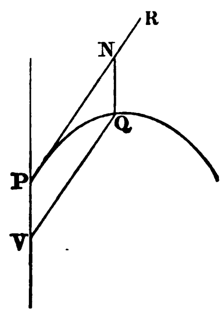
$$x = a + \frac{\frac{3}{2} - \frac{\sqrt{3}}{2}}{3 + 1} \cdot \frac{gt^2}{2}$$

$$= a + \frac{3 - 1.7}{16} gt^2 = a + 2.6t^2, \text{ nearly.}$$

ON PROJECTILES.

37. We now come to the case of motion not rectilinear, and the principal problem to which the formulæ already established are applicable, is that of the motion of a projectile, that is, of a heavy body which has been projected in a direction not vertical. The results which we obtain, it may be observed, are not practically correct, since we omit the consideration of the resistance of the air, and suppose the body to move in a perfect vacuum.

38. *The path of a projectile will be a parabola.*



It is evident that the path will be in one plane: let it lie in the plane of the paper, and let P be the point of projection, PNR the line in which the body is projected, which will manifestly be a tangent to the curve described.

Let V be the velocity of projection, and N the point at which the body would arrive with this velocity in the time t, so that $PN = Vt$.

From N draw NQ vertical, and make $NQ = \frac{gt^2}{2}$; then, since the space which the body would describe in the time t under the action of gravity only is NQ, and if gravity had not acted the body would have been at N, therefore when the body is simultaneously animated by its original velocity V, and that generated by gravity, it will be at Q.

Complete the parallelogram $PVQN$, then

$$PV = NQ = \frac{gt^2}{2},$$

$$QV = PN = Vt;$$

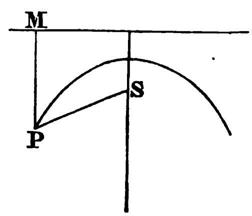
$$\therefore\ QV^2 = V^2t^2 = \frac{2V^2}{g} \cdot PV.$$

But in the parabola $QV^2 = 4SP \cdot PV$. (See Conic Sections, Prop. IX. page 140.) Hence Q lies in a parabola, of which the axis is vertical, and the distance of P from the directrix or focus is $\frac{V^2}{2g}$.

COR. From the point P draw PM perpendicular to the directrix; then

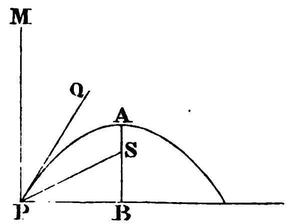
$$V^2 = 2g \cdot SP = 2g \cdot PM.$$

Hence, *the velocity at any point of the parabola is that which would be acquired in falling from the directrix.*



39. *To determine the position of the focus of the parabola.*

Let P be the point of projection, PQ the direction of projection, ASB the axis of the parabola, PM vertical, PB horizontal; and let QPB (which is called the angle of projection) $= \alpha$.



From P draw PS, making the same angle with PQ as PQ makes with PM; then, by a property of the parabola, the point S, in which PS meets the axis, is the focus.

Now $SPB = QPS - QPB = MPQ - QPB$ $\quad$ $SPB = QP$

$$= 90^0 - a - a = 90^0 - 2a; \qquad = \alpha - (90^\circ - \alpha) = 2$$

$$\therefore \quad PB = SP \sin 2a, \qquad \therefore\ PB = SP \sin 2\alpha$$

$$SB = SP \cos 2a, \qquad SB = -SP \cos 2\alpha$$

$$\text{but } SP = \frac{V^2}{2g}; \quad \therefore \; PB = \frac{V^2}{2g} \sin 2a, \qquad \therefore\ PB = \frac{V^2}{2g} \sin 2\alpha$$

$$SB = \frac{V^2}{2g} \cos 2a. \qquad SB = -\frac{V^2}{2g} \cos 2\alpha$$

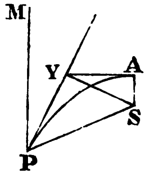
The lines PB, SB, determine the position of S; for we have only to measure the distance PB horizontally, and BS vertically, and we shall determine S.

40. *To determine the greatest height to which the projectile will rise.*

The velocity of projection may be supposed to be resolved into two velocities, one horizontal $= V \cos a$, the other vertical $= V \sin a$. The vertical motion of the body will be the same as if it were projected vertically with this latter velocity; hence, if h be the height required, we have

$$h = \frac{V^2 \sin^2 a}{2g}.$$

41. *To find the latus rectum of the parabola.*



Draw the perpendicular SY on the tangent PY, which is the line of projection; let A be the vertex then AY is a tangent at the vertex, and therefore horizontal. (See Conics, Prop. VI. page 137.)

Then $SYA = 90^0 - \alpha$, and $SPY = MPY = 90^0 - \alpha$;

$\therefore$ the latus rectum $= 4AS = 4SY\cos\alpha = 4SP\cos^2\alpha$

$$= \frac{2V^2}{g}\cos^2\alpha.$$

42. *To find the range of the projectile, that is, the distance from the point of projection at which the body meets the horizontal plane through the point of projection.*

The range will evidently be twice the distance of the point of projection from the axis of the parabola. But this distance was shewn in Art. 39 to be $\frac{V^2}{2g}\sin 2\alpha$;

$$\therefore \text{ the range} = \frac{V^2}{g}\sin 2\alpha.$$

Cor. The greatest value which $\sin 2\alpha$ can have is 1, which corresponds to $\alpha = 45^0$; hence, the range of a projectile will be greatest when the angle of projection is 45^0.

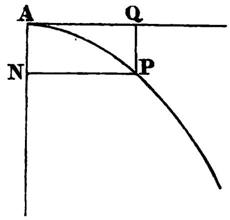
43. The preceding are the principal propositions respecting projectiles. We shall now give a few examples, which may be multiplied to any extent.

Ex. 1. A body is projected horizontally with a velocity of 4 feet per second, to find the latus rectum of the parabola described.

The general expression for the latus rectum is $\frac{2V^2}{g}\cos^2\alpha$; and in this case $\alpha = 0$, and $V = 4$;

$$\therefore \text{ the latus rectum} = \frac{16}{16.1} = 1 \text{ foot, nearly.}$$

When the body is projected horizontally, it is manifest that the point of projection is the vertex of the parabola, and the general investigation of the path becomes much simplified.



Let AQ be the direction of projection, and let Q be the point, at which the body would arrive from A in the time t with a velocity V; $\therefore\ AQ = Vt$.

Again, take $QP = \dfrac{gt^2}{2}$, so that QP is the distance through which the body would fall in time t under the action of gravity only: then P is the actual place of the body at the time t.

Complete the rectangle $ANPQ$; then

$$PN = Vt,$$

$$AN = \frac{gt^2}{2};$$

$$\therefore\ PN^2 = V^2t^2 = \frac{2V^2}{g}\,.\,AN;$$

but in the parabola $PN^2 = 4AS\,.\,AN$ (See Conics, Prop. v. page 137.)

$$\therefore\ 4AS = \frac{2V^2}{g}.$$

Ex. 2. A body is projected at an angle of 30^0, with the velocity which it would acquire in falling through 5 feet; find the range.

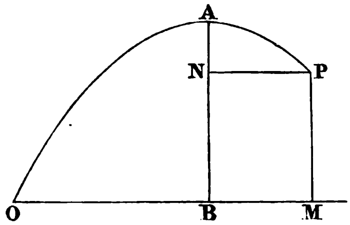
In general, the range $= \dfrac{V^2}{g}\sin 2\alpha$,

and in this case $V^2 = 2g \times 5$;

$$\therefore\ \frac{V^2}{g} = 10,$$

$$\text{and the range} = 10 \sin 60^0 = 5\sqrt{3} \text{ feet.}$$

Ex. 3. To find the relation between the velocity and angle of projection, in order that the projectile may strike a given point.



Let O be the point of projection, OM horizontal, and MP vertical; let P be the given point, then it will be given provided we know OM and MP: let $OM = h$, $MP = k$. Draw AB the axis of the parabola, and PN perpendicular to AB.

Then we have proved, (Arts 39, 40) that $OB = \dfrac{V^2 \sin 2\alpha}{2g}$

$$\text{and } AB = \frac{V^2 \sin^2 \alpha}{2g};$$

$$\therefore \quad AN = \frac{V^2 \sin^2 \alpha}{2g} - k, \quad PN = h - \frac{V^2 \sin 2\alpha}{2g};$$

but the latus rectum $= \dfrac{2V^2}{g} \cos^2 \alpha$;

$$\therefore \quad PN^2 = \frac{2V^2}{g} \cos^2 \alpha \,.\, AN;$$

$$\text{or} \quad \left(h - \frac{V^2 \sin 2\alpha}{2g}\right)^2 = \frac{2V^2}{g} \cos^2 \alpha \left(\frac{V^2 \sin^2 \alpha}{2g} - k\right);$$

$$\text{or} \quad h^2 - 2h\frac{V^2}{g} \sin \alpha \cos \alpha + 2k\frac{V^2}{g} \cos^2 \alpha = 0,$$

which is the relation between α and V required.

Suppose, for instance, that $h = 100$ feet, $k = 10$ feet, that $V = 40$; then to find α, we have

$$100^2 - 100\frac{40^2}{16.1} \sin \alpha \cos \alpha + 10\frac{40^2}{16.1} \cos \alpha = 0.$$

$$10000 - 100 \times 100 \sin \alpha \cos \alpha + 10 \times 100 \cos^2 \alpha = 0, \text{ nearly}$$

$$\text{or,}\quad 10 - 10 \sin a \cos a + \cos^2 a = 0,$$

$$10 - 5 \sin 2a + \frac{1 + \cos 2a}{2} = 0,$$

$$21 = 10 \sin 2a - \cos 2a\,;$$

which is an equation for finding a.

MOTION OF A PARTICLE ON A CURVE.

44. When a heavy particle moves on a plane curve, the plane being supposed to be vertical, a portion of the weight of the body will be employed in producing pressure on the curve, and the other portion in producing motion. Hence, in considering the motion of a body on a curve, we suppose the force of gravity at any point to be resolved into two parts, one acting along the *tangent* of the curve, the other along the *normal;* the former is the part which produces motion, and is the only part with which we shall generally be concerned. We have already considered a particular case of motion on a curve, when we treated of a body falling down an inclined plane; but in that case the inclination of the plane being always constant, the force producing motion was uniform, whereas in the more general problem of motion on any curve the intensity of the force varies from point to point. Consequently the formulæ which we have hitherto used are inapplicable, and we shall be obliged to content ourselves, in this place, with one general proposition, which we can demonstrate by general reasoning. We shall recur to the subject hereafter.

45. *When a heavy body falls down the surface of a smooth curve, the velocity at any point is that due to the vertical height through which the body has fallen.*

In the first place we observe, that the pressure of the curve is always perpendicular to the direction of the motion, and therefore destroys no velocity.

In the next place, the proposition is true for an inclined plane, as has been shewn. (See Page 236.)

And, lastly, we may consider the curve to be a succession of inclined planes, the planes being indefinitely short in length and great in number, and the proposition being true for each will be true for all, and therefore for the curve. For though it may be argued, that there is always a loss of velocity in passing from one plane to another, yet we know that, when the planes are so far increased in number and diminished in magnitude as to be considered coincident with the curve, this will not be the case, since by our first observation no velocity is lost.

46. There is one point connected with the motion of a body on a curve, of which we can give a general explanation, and on which it is desirable to have distinct notions; and that is, the existence of what is called *centrifugal force.* The explanation is applicable, not only to the motion of a body constrained to move upon a curve line or curve surface, but also to that of a body describing any path not rectilinear under the action of any forces.

When a body, acted upon by any force, moves in the direction in which that force acts, it has a tendency at each moment to proceed only with the velocity which it has at that moment; this follows from the first law of motion, or is the consequence of the *inertia* of the body. But when a body, acted upon by any force, moves transversely to the direction of the force, it has a tendency at each moment, not only to proceed with the velocity with which it is then animated, but also to continue to move in the direction in which it is moving at the instant in question; this also is a consequence of the first law of motion. The motion therefore may be conceived of as though the body were under the action of a force, always tending to make the body leave the path which it actually describes; the force which measures this tendency, is called the *centrifugal force*, a name liable to much objection, because there is in fact no force acting on the body to draw it out of its path, the tendency to leave that path being the result of its own motion only.

Hence if we conceive the forces, acting on a body which describes a plane curve, to be resolved into two, one in the direction of the tangent, the other in the direction of the normal, we may say that the former portion is expended in accelerating or retarding the body's motion, the latter in changing its direction, or (in other words) in counteracting the *centrifugal force.* And if the body be constrained to move on a curve, the normal portion will be employed in counteracting the centrifugal force, and also in producing pressure on the curve. Therefore, when a heavy particle moves upon a curve in a vertical plane, the pressure on the curve is not the resolved part of the weight in the direction of the normal, as would be the case if the particle were at rest, but is greater or less than that resolved part according as the centrifugal force tends to increase or diminish the pressure.

ON THE COLLISION OR IMPACT OF BODIES.

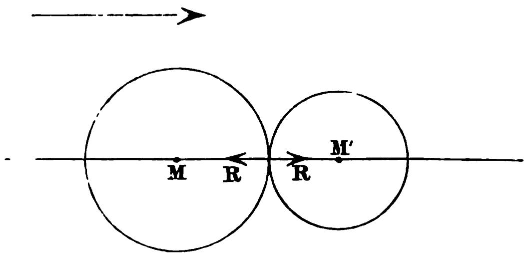
47. In investigating the circumstances of motion which attend the collision of bodies, there are two cases for us to consider; (1) that of *inelastic* bodies, (2) that of *elastic.* We must define the meaning of these terms.

We have already, when speaking of impulsive force, described the nature of the action which takes place when two elastic bodies impinge upon each other; now if when two bodies impinge upon each other, after the compression of their figures due to the impact has ceased, a force of restitution of figure comes into operation, the bodies are called *elastic,* but if there is no such force then they are called *inelastic.* To take examples, ivory and glass are highly elastic substances, a lump of putty or clay is perhaps as free from elasticity as any substance.

Let us consider the impact of inelastic bodies.

48. *Two inelastic balls, moving in the same direction but with different velocities, impinge upon each other; to determine the motion after impact.*

Let M M' be the masses of the two balls, V V' their velocities before impact.



When the bodies impinge, there will be an impulsive pressure between them, which we will call R, and the value of which it must be our business to find. This pressure will of course be equal in magnitude and opposite in its direction upon the two balls, i. e. accelerating one and retarding the other.

The momentum of the ball M before impact is MV, therefore its momentum after impact is $MV - R$, and therefore its velocity $V - \frac{R}{M}$.

Similarly, the velocity of the ball M' after impact is $V' + \frac{R}{M'}$.

But since, by the hypothesis of inelasticity, there is no force after impact to separate the balls, they will proceed with a common velocity;

$$\therefore \; V - \frac{R}{M} = V' + \frac{R}{M'};$$

$$\therefore \; R = \frac{MM'}{M + M'}(V - V'),$$

and the common velocity $= V - \frac{M'}{M + M'}(V - V') = \frac{MV + M'V'}{M + M'}$.

If the balls are moving in *opposite* directions, we have only to write $- V'$ instead of V'.

Cor. If we call v the velocity after impact, we have

$$Mv + M'v = MV + M'V';$$

that is to say, the sum of the momenta of the balls is the same after impact as before it.

49. We can now proceed to the problem of elastic bodies. We may consider the impact as consisting of two parts, viz. during the compression of the bodies, and during the restitution of their forms: as long as compression continues, the problem is precisely the same as if the bodies were inelastic, and if we call the impulsive pressure between them during compression R, the value of R will be that already found on the supposition of the bodies being inelastic; for, though the bodies do not, for any sensible time after impact, move with the same velocity, yet during that very short time in which the compression takes place they do so; hence the *force of compression* is already determined. When the restitution of force takes place, a new force is brought into action, which we shall distinguish as the *force of restitution*, and shall call R'; to determine R' we must have recourse to experiment, and it is found that the ratio of R' to R is independent of the velocity of the bodies, and dependent only on the nature of the substances of which they are composed. So that, if we make $R' = eR$, we may consider e to be a known quantity, since in any given example, if the substance of the bodies is given, the value of e may be found from experiment, or by reference to tables of elasticity. The quantity e is called the *modulus of elasticity*; it is always some quantity less than 1; the limiting case of e being actually $= 1$ is that of *perfect elasticity*, but there is no such case in nature.

50. *Two elastic balls moving in the same direction, but with different velocities, impinge upon one another; to find the velocities after impact.*

Let M M' be the masses of the bodies.

V V' their velocities before impact.

v v' .... after ....

RR' the forces of compression and restitution respectively, so that the whole impulsive force between the balls $= R + R'$ $= R(1 + e)$, where e is the modulus of elasticity.

We may find R on the supposition of the bodies being inelastic; hence by our previous investigation, (Art. 47.)

$$R = \frac{MM'}{M+M'}(V - V');$$

$$\therefore v = V - \frac{R+R'}{M} = V - (1+e)\frac{R}{M} = V - (1+e)\frac{M'}{M+M'}(V-V'),$$

$$v' = V' + \frac{R+R'}{M'} = V' + (1+e)\frac{R}{M'} = V' + (1+e)\frac{M}{M+M'}(V-V').$$

If the balls are moving in opposite directions we must change the sign of V', as in Art. 48.

51. *An elastic ball impinges directly upon a fixed plane; to find the velocity after impact.*

Let V be ball's velocity before impact,

v .... after ......

RR' the forces of compression and restitution,

e the modulus of elasticity.

Then, to find R, we suppose the body inelastic; but in this case there would be no velocity after impact, since the plane is fixed;

$$\therefore V - \frac{R}{M} = 0, \text{ or } R = MV;$$

$$\therefore R + R' = (1+e)MV,$$

$$\text{and } v = V - (1+e)V = -eV.$$

Hence the ball's velocity will be diminished in the ratio of $1 : e$. The negative sign indicates that the motion after impact must be in the opposite direction to that before impact, which must manifestly be the case.

52. By *oblique* impact we intend to express those cases of impact, in which the direction of the velocity does not coincide with the direction of the mutual impulsive pressure.

53. ***A body impinges upon a fixed plane, in the direction of a line making a given angle with the normal to the plane; to determine the motion after impact.***

Let V be the velocity before impact, a the angle which its direction makes with the normal to the plane: v, θ similar quantities after impact. The rest of the notation as before.

We may suppose the velocity V to be resolved into two velocities, one parallel to the plane ($V \sin a$), the other perpendicular to it ($V \cos a$); the former will not be altered by the impact, the latter may be treated as in the case of direct impact and will therefore be diminished in the ratio of $1 : e$. The resolved parts of the velocity after impact, parallel and perpendicular to the plane, are $v \sin \theta$ and $v \cos \theta$ respectively; hence we shall have,

$$v \sin \theta = V \sin a,$$

$$v \cos \theta = -eV \cos a;$$

$$\therefore \cot \theta = -e \cot a,$$

$$\text{and } v^2 = V^2 (\sin^2 a + e^2 \cos^2 a),$$

which equations determine θ and v.

Cor. If the elasticity be perfect, or $e = 1$, we shall have

$$\cot \theta = -\cot a,$$

$$\text{or } \theta = -a,$$

$$\text{and } v^2 = V^2,$$

$$\text{or } v = V.$$

The interpretation of these results is, that the ball will rebound from the plane with a velocity equal to that of incidence, and in a direction making an angle with the normal equal to the angle of incidence, but on the opposite side of the normal.

54. We shall subjoin a few examples of impact.

Ex. 1. *A perfectly elastic ball impinges directly upon another, and this upon a third; compare the velocity communi-*

cated to the third, with that which would have been communicated if the first had impinged upon it.

Let $M\ M'\ M''$ be the masses of the balls, V the velocity of M before impact.

Let R be the force of compression, then we should find, by an investigation such as that in Art. 47, that

$$R = \frac{MM'}{M+M'} V;$$

$\therefore$ the velocity of R' after impact $= \frac{2R}{M'} = \frac{2M}{M+M'} V$, (since $e = 1$).

In like manner, the velocity communicated to M''

$$= \frac{2M'}{M' + M''} \cdot \frac{2M}{M+M'} V.$$

But the velocity, which would have been communicated if M had impinged upon M''

$$= \frac{2M}{M + M''} V;$$

$$\therefore \text{ the ratio required is } \frac{2M'(M + M'')}{(M' + M'')(M + M')}.$$

Ex. 2. *In the direct impact of perfectly elastic bodies, the sum of the masses of the bodies multiplied each by the square of its velocity is the same before and after impact.*

Let MM' be the masses of the bodies, VV' their respective velocities before, and vv' after, impact.

Then, we have seen (Art. 50), that

$$v = V - \frac{2M'}{M+M'}(V - V'),$$

$$v' = V' + \frac{2M}{M+M'}(V - V'),$$

since $e = 1$;

$$\therefore\ v - v' = V - V' - 2(V - V')$$

$$= -(V - V'),$$

$$\text{or}\quad v + V = v' + V' \ldots\ldots\ldots\ldots\ldots (1).$$

Again,

$$Mv + M'v' = MV + M'V',$$

$$\text{or}\quad M(v - V) = -M'(v' - V') \ldots\ldots\ldots (2).$$

Multiplying together (1) and (2), we have

$$M(v^2 - V^2) = -M'(v'^2 - V'^2),$$

$$\text{or,}\quad Mv^2 + M'^2 v'^2 = MV^2 + M'V'^2.$$

The mass of a body multiplied by the square of its velocity is called its *Vis Viva*; hence it appears, that when the elasticity is perfect, the sum of the *Vis Viva* of two impinging bodies is not altered by impact.

Ex. 3. *In the collision of imperfectly elastic bodies, Vis Viva is lost by the impact.*

In this case we have

$$v = V - (1 + e)\frac{M'}{M + M'}(V - V'),$$

$$v' = V' + (1 + e)\frac{M}{M + M'}(V - V');$$

$$\therefore\ Mv + M'v' = MV + M'V';$$

also $\quad v - v' = V - V' - (1 + e)(V - V') = -e(V - V');$

$$\therefore\ (Mv + M'v')^2 = (MV + M'V')^2,$$

and $\quad MM'(v - v')^2 = MM'e^2(V - V')^2$

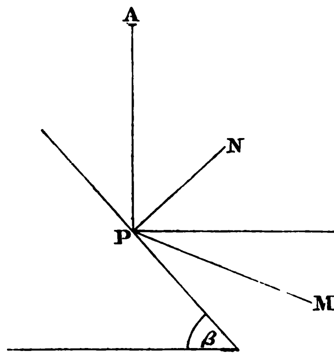
$$= MM'(V - V')^2 - (1 - e^2)MM'(V - V')^2$$

$\therefore$ by addition, $(M+M')(Mv^2+M'v'^2)=(M+M')(MV^2+M'V'^2)$
$$- (1 - e^2)\, MM'\, (V - V')^2,$$

or $Mv^2 + M'v'^2 = MV^2 + M'V'^2 - (1 - e^2)\dfrac{MM'}{M + M'}(V - V')^2$;

which proves the proposition, since e is less than 1.

Ex. 4. *An elastic ball falls from a given height upon a given inclined plane; to find the latus rectum of the parabola described after the rebound.*



Let β be the angle of the plane: A the point from which the body falls; P the point of impact, $AP = h$. Let PN be normal to the plane, PB horizontal, and PM the direction in which the ball goes off after impact, $BPM = \theta$.

Then the velocity with which the body reaches $P = \sqrt{2gh}$, and the angle $APN = \beta$, hence if V be the velocity with which the ball leaves the plane,

$$V^2 = 2gh\,(\sin^2\beta + e^2\cos^2\beta). \quad \text{(Art. 53.)}$$

Again, $NPM = 90^0 - \beta + \theta$, hence we have, (by the same article,)

$$\tan(\beta - \theta) = e\cot\beta.$$

The latus rectum of the parabola described

$$= \frac{2V^2}{g}\cos^2\theta. \quad \text{(Art. 41.)}$$

And we have

$$\frac{\tan\beta - \tan\theta}{1 + \tan\beta\tan\theta} = \frac{e}{\tan\beta},$$

$$\tan^2\beta - \tan\theta\tan\beta = e + e\tan\beta\tan\theta;$$

$$\therefore\ \tan\theta = \frac{\tan^2\beta - e}{\tan\beta(1+e)},$$

and $$\cos^2\theta = \frac{1}{1+\tan^2\theta} = \frac{\tan^2\beta(1+e)^2}{\tan^2\beta(1+e)^2 + (\tan^2\beta - e)^2}$$

$$= \frac{\tan^2\beta(1+e)^2}{\sec^2\beta(\tan^2\beta + e^2)} = \sin^2\beta\,\frac{(1+e)^2}{\tan^2\beta + e^2};$$

$$\therefore\ \frac{2V^2}{g}\cos^2\theta = 4h(\sin^2\beta + e^2\cos^2\beta)\sin^2\beta\,\frac{(1+e)^2}{\tan^2\beta + e^2}$$

$$= 4h\sin^2\beta\cos^2\beta(1+e)^2$$

$$= h\sin^2 2\beta(1+e)^2.$$

NEWTON.

SECTION I.

[DIGRESSION CONCERNING THE CURVATURE OF CURVE LINES.]

SECTION II.

SECTION III.

[APPENDIX TO SECTION II,

CONTAINING THE THEORY OF CYCLOIDAL OSCILLATIONS.]

NEWTON.

INTRODUCTORY REMARKS.

1. It may perhaps assist the student towards the right understanding of those extracts from Newton's Principia which follow, if we preface them with a few remarks respecting their general purpose.

2. The ordinary processes of geometry and trigonometry are sufficient for the mensuration, and for the discussion of the properties, of straight lines and figures bounded by straight lines; but these methods fail when we come to the consideration of curve lines and figures bounded by curves. We require then some method, which shall enable us to extend our calculations to this more difficult case, and such a method is propounded and developed by Newton : he considers that although a figure inclosed by a curve line is not a polygon, yet a polygon may be made to approach as near as we please to such a figure by increasing the number and diminishing the length of its sides; according to a common phraseology, a figure inclosed by a curve line may be regarded as the *limit* of a polygon, and thus we are enabled to extend to the former propositions proved concerning the latter.

We have already, in fact, anticipated this view in one simple case, namely, in determining the area and circumference of a circle, (Trig. Art. 51. page 121.) ; for we deduced those expressions by observing the values to which the area and circumference of a regular polygon continually approximated, when the number of the sides was increased and their length diminished indefinitely.

We also anticipated the principle, when we regarded the tangent of a curve as the line drawn through two points in the curve, when one of those points is made to move up

indefinitely near to the other; in other words, we regarded the tangent as the limiting position of the secant, when the points through which the secant is drawn coalesce. (Conics, page 131.)

3. Having used the term *limit*, let us endeavour to define it strictly.

Def. The limit of the value of a quantity is the value to which the quantity continually approaches, (though it never reaches it,) when any element on which the quantity depends is indefinitely increased or diminished.

The limiting position of a line would be similarly defined. Hence it appears, that we do not assert that a quantity ever is equal to its limit, but only that it continually approximates to it. Thus we do not say that a circle is a polygon, or a tangent a secant, but only that a polygon continually approximates to a circle as the number of its sides is increased and their length diminished, and that a secant continually approaches to a tangent as the points of section approach each other.

4. The difficulty, which we have pointed out as existing in the application of mathematics to Geometry, exists also in the application of them to Mechanics. Thus we have seen, that in our treatise on Dynamics we were restricted to the consideration of uniform force, because we had no calculus which enabled us to treat of force varying from one moment to another. We may however consider, that if we suppose the case of a number of impulses, and suppose these impulses to be indefinite in number but also indefinitely small in intensity, we shall have a hypothetical system of impulses approximating as near as we please to the character of continuous varying force; in other words, a continuous varying force may be regarded as the *limit* of a system of impulses.

5. The method of calculation, which Newton has founded on this idea of a *limit*, and which he has developed in the first section of his Principia, he calls "The method of prime

and ultimate ratios." The propriety of the name will be seen by considering an example.

Let us suppose PQ to be a portion of a curve, PT the tangent to it at P, TQ perpendicular to PT: then we may consider the curve PQ to be traced out by a point, moving according to some given law, and PT to be traced out by a point, which moves in the direction in which the former point was moving at P. Now it will be proved, that the limit of the ratio of the lines PT and PQ, when TQ is drawn indefinitely near to P, is one of equality; hence, if we regard PQ and PT as described by two points beginning to move from P, we may speak of their *nascent* state, and say that their *prime* ratio (that is, the ratio which they have *at first*,) is one of equality; or, on the other hand, we may suppose PQ and PT to be continually diminished by the approach of TQ to P, and then we may speak of their *vanishing* state and say, that their *ultimate* ratio is one of equality. *Prime* and *ultimate* therefore are, in general, expressions for the same thing, contemplated from two different points of view.

6. It may be well to observe, that when Newton speaks of two quantities being *ultimately equal*, he does not mean that they ever are really equal, but that they are tending to the same limit; thus, to take an algebraical illustration, according to Newton's phrase, $a + x$ and $a + 2x$ are ultimately equal when x is indefinitely diminished, because both tend to the same limit, viz. the quantity a.

And, in like manner, in the example adduced in the preceding article, Newton would speak of PT and PQ being ultimately equal; not asserting thereby that those lines are ever really equal, but only that they constantly tend to equality, and that the difference between their ratio and unity diminishes indefinitely as the line TQ approaches P.

NOTE. In the following version of the three sections, some demonstrations and propositions have been introduced which are not found in the Principia: all such interpolations are marked by being inclosed in parentheses.

SECTION I.

ON THE METHOD OF PRIME AND ULTIMATE RATIOS.

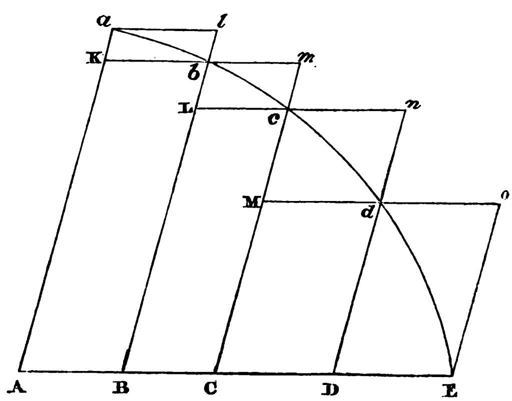
LEMMA I.

Quantities and the ratios of quantities, which tend constantly to equality, and may be made to approximate to each other by less than any assignable difference, become ultimately equal.

For if not, let them become ultimately unequal, and their difference be ultimately *D*. Therefore they cannot approximate to each other by less than the difference *D*, and this is contrary to the hypothesis, which is, that they *may* approximate by less than any assignable difference. Wherefore they do not become ultimately unequal, that is, they become ultimately equal. Q.E.D.

LEMMA II.

If in any figure AaE, *bounded by the straight lines* Aa, AE *and the curve* acE, *there be inscribed any number of parallelograms* Ab, Bc, Cd...... *on equal bases* AB, BC

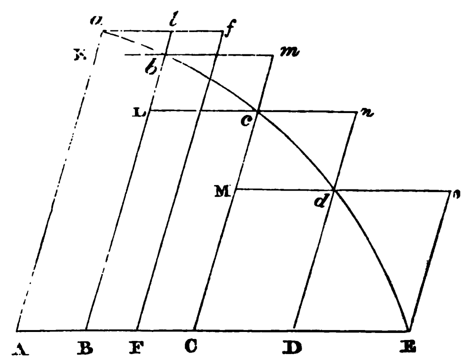


CD...... *and sides* Bb, Cc...... *parallel to the side of the figure* Aa, *and the parallelograms* aKbl, bLcm, cMdn... *be completed: then, if the breadth of these parallelograms be diminished and their number increased indefinitely, the ultimate ratios of the inscribed figure* AKbLcM......*the circumscribed figure* AalbmcndoE, *and the curvilinear figure* AabcdE, *will be ratios of equality.*

For the difference of the inscribed and circumscribed figures is the sum of the parallelograms *Kl*, *Lm*, *Mn*, *Do*......, that is, (since the bases are all equal) the parallelogram *AalB*. But this parallelogram, by diminishing its breadth indefinitely, may be made less than any assignable quantity. Therefore, by Lemma I., the inscribed and circumscribed figures, and *à fortiori* the curvilinear figure which is intermediate to the two, become ultimately equal. Q.E.D.

LEMMA III.

The same ultimate ratios are also ratios of equality, when the breadths of the parallelograms AB, BC, CD,...... *are unequal, and all are diminished indefinitely.*



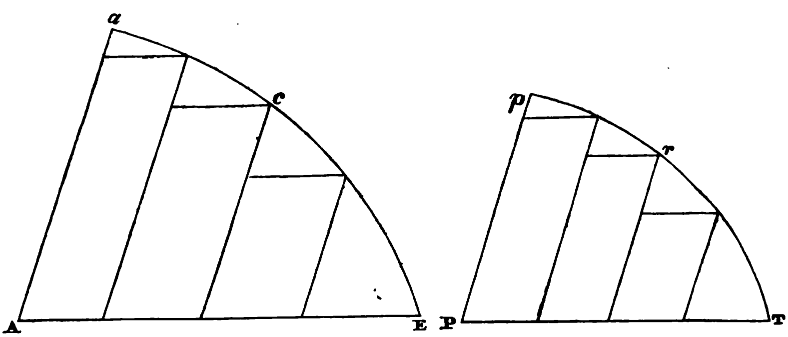
For let *AF* be equal to the greatest breadth, and complete the parallelogram *AafF*. Then this parallelogram will be greater than the difference between the inscribed and circumscribed figures; but, when its breadth is diminished indefinitely, it will become less than any assignable quantity, and

therefore *á fortiori* the difference between the inscribed and circumscribed figures will be leſs than any assignable quantity. Hence, as in the preceding Lemma, the ultimate ratios of the inscribed, the circumscribed, and the curvilinear figure, will be ratios of equality. Q.E.D.

Cor. If ab, bc, cd...... be joined, the rectilinear figure so formed coincides ultimately with the curve line $abcdE$, and the area $Aabcd$...... with the area AaE.

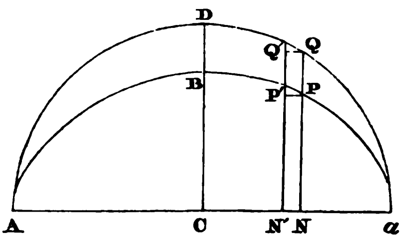
Lemma IV.

If in two figures AacE, PprT, *are inscribed two series of parallelograms, (as in the preceding Lemmas,) the number in the two series being the same, and if when the breadths of the parallelograms are diminished and their number increased indefinitely, the ultimate ratios of the parallelograms in one figure to those in the other each to each are all the same; then are the figures* AacE, PprT *in that same ratio.*



For as the parallelograms are each to each, so (*componendo*) is the sum of all in one figure to the sum of all in the other, and therefore the figure $AacE$ to the figure $PprT$; for, by the preceding Lemma, the ultimate ratio of these figures to the series of inscribed parallelograms is a ratio of equality. Q.E.D.

[Cor. By means of this Lemma we may find the area of an ellipse.



For if ABa be the ellipse, ADa the auxiliary circle, and we describe in these a series of parallelograms on equal bases, such as $PP'N'N$, $QQ'N'N$, these parallelograms are to each other as $PN : QN$, or as $BC : AC$, (Conics, Prop. vi. page 148.)

Therefore

$$\text{area of ellipse} : \text{area of circle} :: BC : AC,$$

$$\text{or, area of ellipse} = \frac{BC}{AC} \cdot \pi AC^2$$

$$= \pi AC \cdot BC.]$$

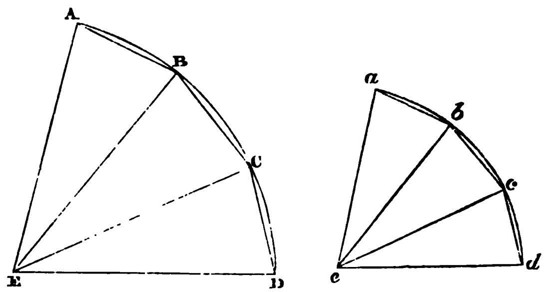
[Def. One curvilinear figure is said to be similar to another, when any rectilinear figure being inscribed in the first, a similar rectilinear figure may be inscribed in the other.

In other words, similar curvilinear figures are the *limit* of similar rectilinear figures, the sides of which have been indefinitely increased in number and diminished in length.]

Lemma V.

The homologous sides of similar curvilinear figures are proportional, and their areas are in the duplicate ratio of the sides.

[Let AED, aed be two similar curvilinear figures, of which the sides AE, ED, AD are homologous to ae, ed, ad, respectively; then, by definition, if $ABCDE$ be a polygon inscribed in one, a similar polygon $abcde$ may be inscribed in



the other. Join $EB, EC \ldots\ldots eb, ec \ldots\ldots$, dividing the polygons into the same number of similar triangles;

$$\therefore\ AB : ab :: AE : ae,$$

$$\text{similarly } BC : bc :: BE : be :: AE : ae,$$

$$CD : cd :: AE : ae,$$

$$\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots$$

$\therefore$ *componendo,*

$$AB + BC + CD + \ldots : ab + bc + cd \ldots :: AE : ae.$$

Now this, being always true, will be true when the number of sides is increased and their lengths diminished indefinitely; but, in this case, the rectilinear figure $ABCD \ldots$ becomes ultimately equal to the curve line AD, and $abcd \ldots\ldots$ to ad;

$$\therefore\ AD : ad :: AE : ae :: ED : ed.$$

Again,

$$\text{polygon } EABCD : \text{polygon } eabcd :: AE^2 : ae^2,$$

and this being always true will be true in the limit as before; therefore, (Lemma III. Cor.)

$$\text{curvil}^r.\text{ figure } AED : \text{curvil}^r.\text{ figure } aed :: AE^2 : ae^2$$
$$:: AD^2 : ad^2$$
$$:: ED^2 : ed^2.$$

Q. E. D.

Cor. If AED, aed are similar figures, and EC, ec equally inclined to ED, ed, then $EC : ec :: ED : ed$.]

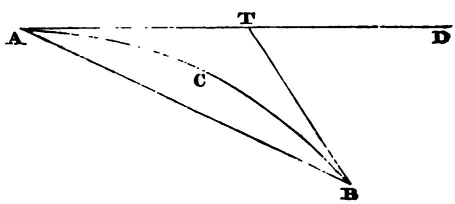
[A *curve* may be conceived as being traced by a moving point, the direction of the motion of which is continually changing.

The *tangent* at any point of a curve, thus considered, is the straight line, in which the generating point would move, if instead of changing the direction of its motion it moved on in the direction which it had at the given point.

A curve is said to be one of *continued curvature*, when the change of direction is not abrupt, but gradual; that is, if ACB (fig. Lemma VI.) be an arc of continued curvature, AD a tangent at A, and BT a tangent at B, then as the point B moves to A the angle BTD which determines the direction of its motion diminishes, not abruptly, but gradually, and ultimately vanishes.]

Lemma VI.

If any arc of continued curvature ACB *be subtended by the chord* AB, *and have the tangent* ATD *at* A; *then if the point* B *move up to* A, *the angle* BAD *will diminish indefinitely and ultimately vanish**.

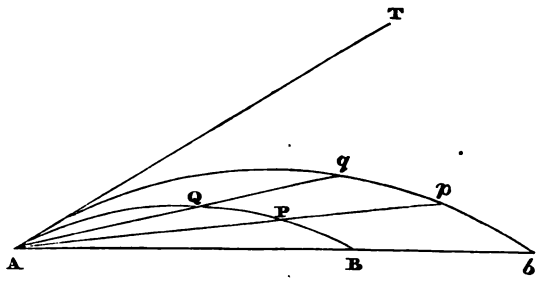


[Draw the tangent BT at B; then, since the curvature is continued, the angle BTD continually diminishes as B ap-

* It will be easily seen, that the mode of viewing the tangent to a curve, which has been here adopted, coincides with that according to which the tangent is considered as the limiting position of the secant, (see page 131); for, since the angle BAD ultimately vanishes, the secant AB ultimately coincides with the tangent AD.

proaches A, and ultimately vanishes; therefore *à fortiori* the angle BAT, which is less than BTD, continually diminishes and ultimately vanishes. Q.E.D.

COR. Similar conterminous arcs, which have their chords coincident, have a common tangent.



Let the similar conterminous arcs APB, Apb, have their chords AB, Ab coincident, and let APp, AQq be any other coincident chords; then since the curves are similar,

$$AQ : Aq :: AB : Ab :: AP : Ap.$$

Hence the arcs AQP, Aqp are similar, and therefore, if P move up to A, the arcs AP, Ap being always similar will vanish together, and the chord APp in its ultimate position will be a tangent to both.

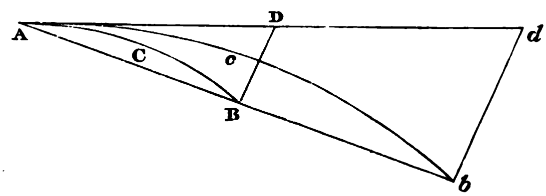
DEF. The *subtense* of an arc is a straight line, drawn from one extremity of the arc to meet at a finite angle, the tangent to the arc, at its other extremity.

OBS. The following three Lemmas involve a common principle, which it may be well to endeavour to explain. The purpose of each Lemma is to discover the ultimate value of a ratio, both terms of which become in the limit evanescent, and the difficulty consists in determining this value geometrically. The artifice made use of by Newton, is this; he substitutes for the ratio, the ultimate value of which is to be determined, another ratio, which is such, that it is always equal to the given ratio, but yet that its terms become finite and not evanes-

cent in the limit; the difficulty therefore, just now alluded to, does not enter into the determination of the ultimate value of this subsidiary ratio, which being found, the ultimate value of the given ratio is also known, being equal to it.]

LEMMA VII.

If BD *be a subtense of the arc* ACB *of continued curvature, and* B *move up to* A, *then will the ultimate ratio of the arc* ACB, *the chord* AB, *and the tangent* AD *be a ratio of equality.*



Let AD be produced to some fixed* point d, and as B moves up to A, suppose db always drawn through d parallel to DB to meet AB produced in b. Also on Ab suppose an arc Acb to be described always similar to the arc ACB, and having therefore ADd for its tangent.

Then, by similar figures, we shall always have,

$$AB : ACB : AD :: Ab : Acb : Ad;$$

and since this proposition is always true, it is true in the limit when B has moved up to A.

But, in this case, the angle bAd vanishes, and therefore the point b coincides with d, and the lines Ab, Ad, and therefore Acb which lies between them, are equal.

Hence also the arc ACB, the chord AB, and the tangent

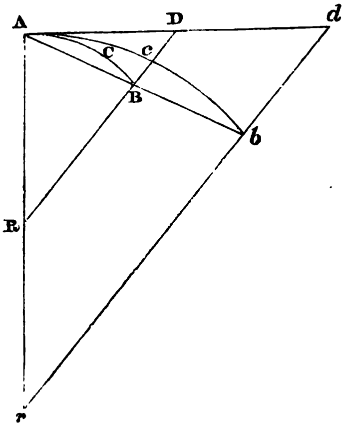
* [For simplicity's sake the point d is spoken of as a *fixed* point, but this condition is not necessary to the proof: the only necessary condition is, that Ad should always be *finite*. A similar observation applies to the next two Lemmas.]

AD, which are always in the same proportion as *Acb*, *Ab*, and *Ad*, are ultimately equal*. Q.E.D.

COR. Hence in all reasonings concerning ultimate ratios, the arc, chord, and tangent may be used indifferently one for another.

LEMMA VIII.

If two straight lines AR, BR, *make with the arc* ACB, *the chord* AB, *and the tangent* AD, *the three triangles* RAB, RACB, RAD; *then when the point* B *moves up to* A, *the three triangles will be ultimately similar and equal.*



Let *AD* be produced to some fixed point *d*, and as *B* moves up to *A* suppose *dbr* always drawn through *d* parallel

* [Hence it may be concluded, that if we use the *circular measure* (see Trigonometry, page 125, Art. 54.) of an angle θ, then the numerical values of θ, $\sin\theta$, $\tan\theta$, and chd θ are ultimately the same.

For, if we consider *ACB* to be an arc of a circle, we have at once, from the Lemma,

$$\text{chd } \theta = \theta \text{ ultimately;}$$

$$\therefore \sin\frac{\theta}{2} = \frac{\theta}{2}, \text{ or } \sin\theta = \theta,$$

$$\text{and } \tan\theta = \frac{\sin\theta}{\cos\theta} = \theta, \text{ since } \cos\theta = 1 \text{ ultimately.}$$

Hence, when θ is very small, θ, $\sin\theta$, $\tan\theta$, and chd θ may, for all practical purposes, be used indiscriminately.]

to DBR to meet AB produced in b, and AR produced in r. Also, on Ab suppose an arc Acb to be described always similar to the arc ACB, and having therefore ADd for its tangent.

Then, by similar figures, we shall always have,

$$RAB : RACB : RAD :: rAb : rAcb : rAd.$$

And since this proportion is always true, it is true in the limit when B has moved up to A.

But, in this case, the angle bAd vanishes, and therefore the point b coincides with d, and Ab with Ad; and the triangles rAb, rAd, and therefore $rAcb$ which is intermediate to them, are similar and equal.

Hence also, the triangles RAB, $RACB$, RAD which are always similar to, and in the same proportion as, rAb, $rAcb$, rAd, are ultimately similar and equal. Q.E.D.

Cor. Hence in all reasonings concerning ultimate ratios, the three triangles aforesaid may be used indifferently for one another.

Lemma IX.

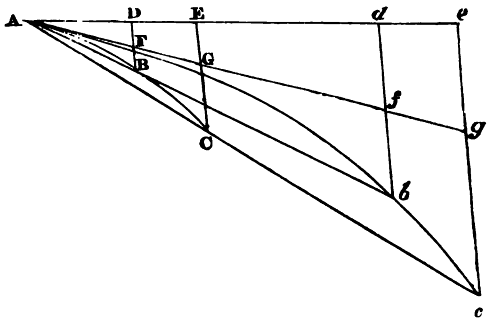
If the straight line AE *and curve* ABC, *given in position, cut each other in a finite angle at* A, *and the lines* BD, CE *be drawn, meeting the line* AE *in any other finite angle, and the curve in* B *and* C; *then, if the points* B *and* C *move up to* A, *the curvilinear triangles* ABD, ACE *will be ultimately in the duplicate ratio of their sides.*

Let AE be produced to some fixed point e, and take Ad such that

$$Ad : Ae :: AD : AE.$$

Draw db, ec parallel to DB, EC, to meet AB, AC produced in b and c. On Ac describe an arc of a curve similar to ABC, which will pass through b, because by similar figures

$$Ab : Ac :: AB : AC.$$



As the points B, C move up to A, suppose the curve Abc to change its form so as to be always similar to the curve ABC; then the area ABD will always be similar to Abd, and ACE to Ace; hence

$$\text{area } ABD : \text{area } Abd :: AD^2 : Ad^2,$$

$$\text{and area } ACE : \text{area } Ace :: AE^2 : Ae^2;$$

$$\text{but } AD^2 : AE^2 :: Ad^2 :: Ae^2;$$

$$\therefore \text{ area } ABD : \text{area } ACE :: \text{area } Abd : \text{area } Ace,$$

and this, being true always, will be true ultimately when B and C have moved up to A.

But, in this case, if $AFGfg$ be the common tangent to the two arcs at A, the angles bAf, cAg will vanish, and the areas Abd, Ace will be ultimately the areas Afd, Age; but

$$\text{area } Afd : \text{area } Age :: Ad^2 : Ae^2,$$
$$:: AD^2 : AE^2;$$

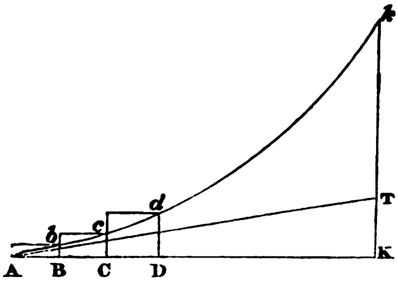
$\therefore$ also, ultimately,

$$\text{area } ABD : \text{area } ACE :: AD^2 : AE^2.$$

Q.E.D.

Lemma X.

The spaces, described from rest by a body under the action of any finite force, are in the beginning of the motion as the squares of the times in which they are described.



Let time be represented by spaces set off along the line AK, and velocity generated by lines perpendicular to AK. And let the time be divided into a number of equal intervals AB, BC, CD, ... ; let Bb, Cc, Dd, ... be the velocities acquired in the times AB, AC, AD...... ; and complete the parallelograms Ab, Bc, Cd......

Suppose the force to act by impulses, which would cause the body to move during the times AB, BC, CD,...... uniformly, with the velocities Bb, Cc, Dd, ... respectively; then the spaces described during the 1st, 2nd, 3rd ...... intervals will be represented by the parallelograms Ab, Bc, Cd, ... and the space described in any given time (AK) by the sum of such parallelograms. But, if we suppose the intervals of time indefinitely decreased in magnitude and increased in number, the series of impulses will constitute a continuous force, and the sum of the parallelograms will (by Lemma II.) be equal to the area AKk.

Hence, if a finite force act during any times AD and AK, we shall have,

space in time AD : space in time AK :: area ADd : area AKk.

Also the angle at which the curve Ak, or the tangent AT, meets the line AK is finite, for since the force is finite the ratio $Kk : AK$ is always finite, and therefore the ratio $KT : AK$ (to which the ratio $Kk : AK$ is ultimately equal) is finite.

Hence, (by Lemma IX.) ultimately,

$$\text{area } ADd : \text{area } AKk :: AD^2 : AK^2;$$

that is, in the beginning of the motion, the spaces described are proportional to the squares of the times of describing them*. Q.E.D.

[Cor. 1. Hence we may obtain an expression for the force (F), under the action of which a body will move from rest, through a given space (S), in a given time (T).

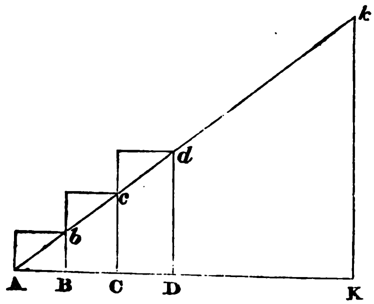
For accelerating force is measured by the velocity which would be generated in a given time divided by the time, the force being supposed uniform throughout the time. (See Dynamics, page 223, Art. 15.) Now, if the force were to be uniform and of the same intensity as at A, the curve Ak would coincide with the tangent AT;

$$\therefore F = \frac{KT}{AK} = \frac{KT.AK}{AK^2} = 2\,\frac{\text{triangle } AKT}{AK^2}$$

$$= 2 \text{ limit } \frac{\text{area } AKk}{AK^2}$$

$$= 2 \text{ limit } \frac{S}{T^2}.$$

* [The same mode of demonstration is applicable to the proposition already proved, (page 229, Art. 29.), namely, that in the case of *uniform* finite force $s = \frac{ft^2}{2}$.



For, let time be represented by spaces set off along the line AK, and velocity generated by lines perpendicular to it, as before. Then, since the velocity is proportional to the time in which it is generated, the points b, c, d...... will be in a straight line; and the space described in the time AK will be represented by the triangle AKk, or by half the rectangle under AK and Kk;

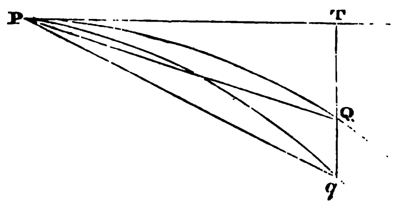
$$\therefore \text{space} = \frac{\text{vel.} \times \text{time}}{2} = \frac{\text{force} \times \overline{\text{time}}|^2}{2},$$

$$\text{or } s = \frac{ft^2}{2}.]$$

COR. 2. The effect produced by F upon the body, is independent of any motion which it may have when F begins to act upon it. Hence, if S be the space through which a force F draws a body, in the time T, from the position which it would have occupied if F had not acted, $F = 2 \text{ limit } \frac{S}{T^2}$.]

[DIGRESSION CONCERNING THE CURVATURE OF CURVE LINES.

1. *On the measure of the curvature of a curve at any point.*



Let PQ, Pq be two curves having the same tangent at P, then the curvatures of these two curves at the point P will be compared, by comparing the rate at which their deflection from the common tangent *begins* to take place. Draw the subtense TQq, and join PQ, Pq; then if TQq were to move parallel to itself up to P, PQ, Pq would ultimately become tangents to the curves PQ, Pq respectively, and the ultimate value of the ratio of the angles TPQ, TPq will therefore measure the ratio of the curvatures of the curves at P.

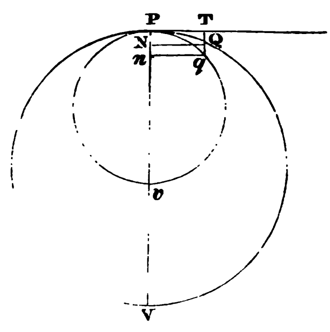
$$\therefore \frac{\text{curvature of } PQ \text{ at } P}{\text{curvature of } Pq \text{ at } P} = \text{limit } \frac{QPT}{qPT} = \text{limit } \frac{\sin QPT}{\sin qPT}$$

$$= \text{limit } \frac{\frac{QT}{PQ} \sin PTQ}{\frac{qT}{Pq} \sin PTq} = \text{limit } \frac{QT}{qT}.$$

2. Another, and perhaps a simpler, way of viewing this proposition is to consider, that if from any point T in the tangent we draw a perpendicular to meet the two curves in Q and q respectively, the deflections from the tangent will be measured by the distances of Q and q from the tangent, that is, by QT and qT. Hence the ratio of the deflections of the two curves from the tangent, in the immediate neighbourhood of P, will be measured by the ultimate value of the ratio $\frac{QT}{qT}$. We have supposed here that QT and qT are drawn perpendicular to the tangent, but the ratio will be the same in the limit at whatever angle they are drawn.

3. The curvature of a circle is the same throughout and depends only on the radius, as we shall shew immediately; hence it is convenient to speak of the curvature of a curve at a proposed point, as being the same as that of a circle of given radius.

4. *The curvatures of two circles are to each other in the inverse ratio of their diameters.*



Let PQV, Pqv be two circles having diameters PV, Pv, and a common tangent PT; from any point T in the tangent draw TQq parallel to PV, and draw the ordinates QN, qn.

Then $\dfrac{\text{curvature of } PQV}{\text{curvature of } Pqv} = \text{limit}\,\dfrac{QT}{qT} = \text{limit}\,\dfrac{PN}{Pn}$

$$= \text{limit}\,\dfrac{\dfrac{NQ^2}{NV}}{\dfrac{nq^2}{nv}} = \text{limit}\,\dfrac{nv}{NV} = \dfrac{Pv}{PV}.$$

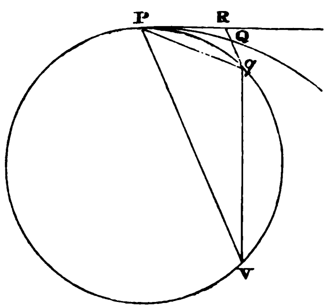
5. Hence, if at any point of a curve we draw a circle, having the same tangent and curvature as the curve has at that point, we may take the reciprocal of its diameter as the measure of the curvature of the curve at that point, and the curvature is said to be finite when the diameter of the circle is finite.

This circle is called *the circle of curvature*, and the radius, diameter, and chord of the circle, are called respectively the *radius*, *diameter*, and *chord of curvature*.

We shall now shew how to calculate the chord and radius of curvature of a curve, and apply the method to the Conic Sections.

6. *If* PqV *be the circle of curvature at any point* P *of a curve* PQ, *and* PV *a chord of the circle drawn in any given direction, then*

$$PV = \text{limit}\,\frac{\text{arc}^2}{\text{subtense parallel to the chord}}.$$



Let RQq be the subtense; join Pq, qV. Then the triangles PVq, PRq are evidently similar,

$$\therefore \; PV = \frac{Pq^2}{Rq}$$

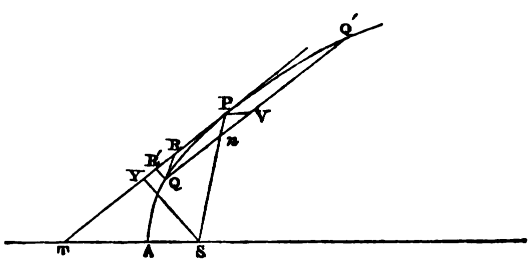
$$= \text{limit}\,\frac{Pq^2}{Rq} = \text{limit}\,\frac{PQ^2}{RQ},$$

since, by hypothesis, limit $\dfrac{RQ}{Rq} = 1$.

Cor. Hence,

$$\text{diameter of curvature} = \text{limit}\,\frac{\text{arc}^2}{\text{subtense perpendicular to tangent}}.$$

7. *To find the chord of curvature through the focus, and the diameter of curvature, at any point of a parabola.*



Draw the tangent PT, QR parallel to SP, QVQ' parallel to PT, and PV parallel to the axis. And let SP, QQ' intersect in n, then $Pn = PV$. (Conics, Prop. II. page 134.)

$$\text{Chord of curvature through } S = \text{limit}\,\frac{PQ^2}{RQ} = \text{limit}\,\frac{Qn^2}{Pn}$$

$$= \text{limit}\,\frac{QV^2}{PV} = 4SP.$$

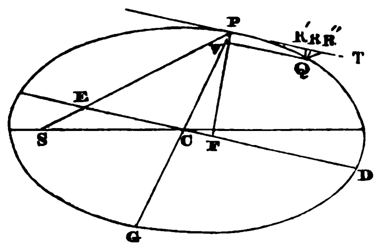
(Conics, Prop. IX. page 140.)

To find the diameter of curvature, draw QR', SY, perpendicular to the tangent: then,

$$\text{diameter of curvature} = \text{limit}\,\frac{PQ^2}{R'Q} = \text{limit}\,\frac{PQ^2}{RQ}\,\frac{1}{\sin SPY}$$

$$= \frac{4SP}{\sin SPY} = 4\,\frac{SP^2}{SY}.$$

8. *To find the chord of curvature through the centre, the diameter of curvature, and the chord of curvature through the focus, at any point of an ellipse.*



Let C be the centre; draw QR parallel to CP, QV to the tangent PT: then, chord through the centre

$$= \text{limit}\,\frac{PQ^2}{RQ} = \text{limit}\,\frac{QV^2}{PV} = \text{limit}\,\frac{CD^2}{CP^2}\,.\,VG$$

(Conics, Prop. VIII. page 150.)

$$= 2\,\frac{CD^2}{CP}\,,\ \text{since}\ VG = 2CP\ \text{ultimately.}$$

Draw PF perpendicular to CD, and QR' to PT; then,

$$\text{diameter of curvature} = \text{limit}\,\frac{PQ^2}{R'Q} = \text{limit}\,\frac{PQ^2}{RQ}\,\frac{1}{\sin QRR'}$$

$$= 2\,\frac{CD^2}{CP}\,\frac{1}{\sin PCD} = 2\,\frac{CD^2}{PF}\,.$$

Again, join SP cutting the conjugate in E, and draw QR'' parallel to it; then,

chord of curvature through the focus

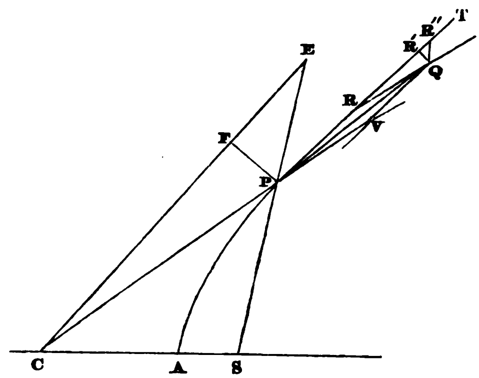
$$= \text{limit}\,\frac{PQ^2}{QR''} = \text{limit}\,\frac{PQ^2}{QR'}\,\sin QR''R'$$

$$= 2\,\frac{CD^2}{PF}\,\sin PEF = 2\,\frac{CD^2}{PE} = 2\,\frac{CD^2}{AC}\,.$$

(Conics, Prop. III. Cor. page 145.)

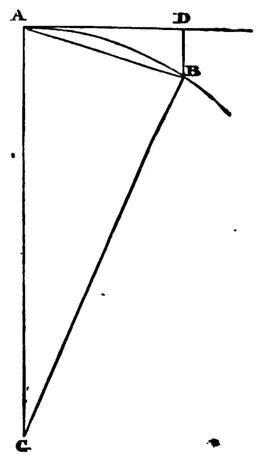
9. *The same for the hyperbola.*

The investigations are the same as for the ellipse; we shall however subjoin a figure.



Obs. The three preceding propositions belong properly to the treatise on Conic Sections, but could not be introduced until the student was familiar with the principles of limits.

10. The following proposition will be required in the succeeding Lemma.



If in the curve AB, AG, BG be drawn perpendicular to the tangent AD and chord AB respectively; then, when B moves up to A, AG will be ultimately the diameter of curvature at A.

Draw BD parallel to AG, then the triangles GAB, ABD are similar;

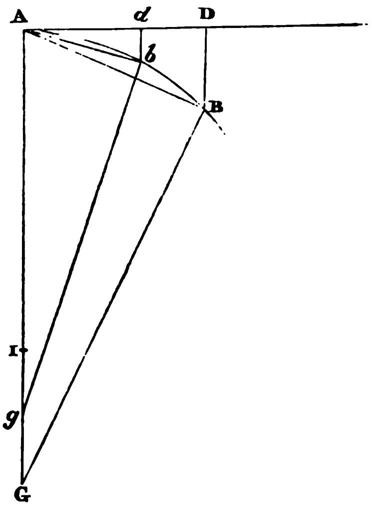
$$\therefore\ AG = \frac{AB^2}{BD};$$

$$\therefore \text{limit } AG = \text{limit} \frac{AB^2}{BD} = \text{limit } \frac{(\text{arc } AB)^2}{BD}$$

$$= \text{diam. of curvature.}]$$

LEMMA XI.

In curves of finite curvature, the subtenses are ultimately in the ratio of the squares of the conterminous arcs.



Let AbB be the curve, having the curvature at A finite.

CASE 1. Let the subtenses bd, BD, be perpendicular to the tangent. Draw bg, BG perpendicular to the chords Ab AB, and let them meet the normal at A in g and G respectively.

Then, when b and B move up to A, g and G will ultimately coincide with I the extremity of the diameter of curvature.

By similar triangles, BAD, AGB, and bAd, Agb,

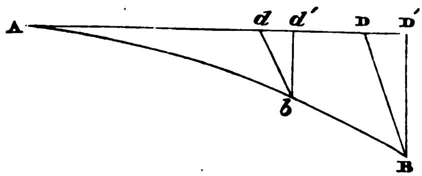
$$BD = \frac{AB^2}{AG}, \quad bd = \frac{Ab^2}{Ag};$$

$$\therefore \; BD : bd :: \frac{AB^2}{AG} : \frac{Ab^2}{Ag}.$$

And, ultimately, $BD : bd :: \dfrac{AB^2}{AI} : \dfrac{Ab^2}{AI}$,

$$:: \overline{\text{arc } AB}\,\rceil^2 : \overline{\text{arc } Ab}\,\rceil^2.$$

CASE 2. Let the subtenses be inclined at any given angle to the tangent. Draw bd', BD', perpendicular to the tangent;



then, by similar triangles, BDD', bdd',

$$BD : bd :: BD' : bd';$$

but, ultimately, $BD' : bd' :: \overline{\text{arc } AB}|^2 : \overline{\text{arc } Ab}|^2$, by Case 1;

$$\therefore \text{ ultimately, } BD : bd :: \overline{\text{arc } AB}|^2 : \overline{\text{arc } Ab}|^2.$$

CASE 3. Suppose the angle D not to be given, but let the lines BD, bd pass through a fixed point, or let B and b approach A according to any other fixed law.

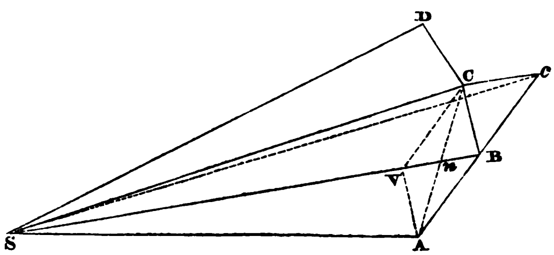
Then, since the angles D and d are formed according to a common law, they will continually approximate to each other as B and b approach A, and will be ultimately equal. Hence this case is reduced to the preceding, and the Lemma is therefore still true.

SECTION II.

ON THE METHOD OF FINDING THE CENTRIPETAL FORCE, UNDER THE ACTION OF WHICH A BODY WILL DESCRIBE A GIVEN ORBIT ABOUT A FIXED CENTRE.

PROP. I. THEOR. I.

The areas described by the lines drawn from the moving body to the fixed centre of force, are all in one plane, and are proportional to the time of describing them.



Let the time be divided into equal parts, and in the first period let the body move with an uniform velocity from A to B. In the second period of time, it would, if acted upon by no force, move in AB produced to c, Bc being $= AB$: and the areas described about the centre S in these two periods, viz. ASB, BSc would be equal; being triangles on equal bases and of equal altitude.

But when the body arrives at B, suppose it to receive a sudden impulse towards S, in consequence of which it moves in the direction BC instead of Bc; then, if we draw cC parallel to BS, the point C will be the actual place of the body at the end of the second period; and the area described will be BSC, which is in the same plane with ASB, because Bc and BS are both in that plane.

Since BSC, BSc are triangles on the same base and between the same parallels, $\therefore BSC = BSc = ASB$.

In like manner, if the body receive an impulse towards S when at C, and D be the place of the body at the end of the third period, we shall have $CSD = BSC = ASB$; and so on, that is, equal areas are described in equal times, and therefore, *componendo*, the sum of the areas described in any given time is proportional to the time, and they all lie in the same plane.

Now suppose the periods of time to be indefinitely diminished in length and increased in number; then the broken line $ABCD$ ... will become ultimately a continuous curve, the series of impulses a continuous central force, and the sum of the areas ASB, BSC, CSD ... the area described by lines drawn from the body to the centre. Hence the areas described, &c. Q.E.D.

COR. 1. The velocity of a body in a central orbit varies inversely as the perpendicular from the centre on the tangent.

For AB, BC, CD, ... are ultimately in the direction of the tangent, and proportional to the velocity, at the points A, B, C... respectively: hence, velocity at $A \propto AB$. But, if we draw a perpendicular p upon AB from S, we have

$$\frac{AB \times p}{2} = \text{area } ASB, \text{ which is constant;}$$

$$\therefore AB \propto \frac{1}{p};$$

$$\text{or, velocity at } A \propto \frac{1}{p}.$$

[We may express this otherwise; let h = twice the area described in a unit of time, and let AB be described in the time t;

$$\therefore \text{ area } ASB = \frac{ht}{2}, \text{ and } AB = vt;$$

$$\therefore vp = h, \text{ or } v = \frac{h}{p}.]$$

COR. 2. If on AB, BC, the chords of two arcs described in equal times, we construct the parallelogram $ABCV$, the diagonal BV will, when the arcs are indefinitely diminished, ultimately, if produced, pass through the centre of force.

COR. 3. The intensity of the central force at B is proportional to the line BV; that is, if $B'V'$ be the line corresponding to BV at some other point B' of the orbit, then, ultimately,

$$\text{force at } B : \text{force at } B' :: BV : B'V'.$$

COR. 4. The forces by which bodies are drawn from their rectilinear motion in curved paths, are proportional to those sagittæ* of arcs described in equal times, the directions of which pass through the centre of force and bisect the chord when those arcs are indefinitely diminished.

For if we join AC, cutting SB in n, Bn will be ultimately one of the sagittæ, but $Bn = \frac{1}{2} BV$, and hence, by the preceding corollary, in the same orbit the force ultimately $\propto Bn$. Also, in different orbits, if we take arcs described in equal times, the sagittæ will measure the effects of the central forces in equal times, i.e. the forces will be proportional to the sagittæ.

PROP. II. THEOR. II.

Every body, which moves in a plane curve, in such a manner that the areas described by lines drawn from it to a fixed point are proportional to the time of describing them, is acted upon by a central force tending to that point.

With the same figure as in the last proposition, let S be the point; and suppose that a body unattracted by any force would describe the space AB in a given interval of time.

Produce AB to c, and make $Bc = AB$; then, if suffered to proceed, the body would, in a second equal interval, arrive

* The *sagitta* of an arc is a line drawn from a point in the chord to a point in the arc.

at c; but at B suppose a sudden impulse communicated, which causes it to move to C, C being such that the triangles SBC, SAB are equal.

Join Cc, Sc; then the triangle $SBC = SAB = SBc$; therefore Cc is parallel to SB, and therefore the impulse at B was in the direction of SB, or tended to S. Similarly, if the body receives at equal intervals of time impulses which make it describe equal triangles in equal times, or a polygonal area proportional to the time, the impulses all tend to S.

The same will be true if we suppose the number of the intervals indefinitely increased and their length diminished, in which case the system of impulses becomes a continuous force, and the polygonal area curvilinear. Hence every body, &c. Q.E.D.

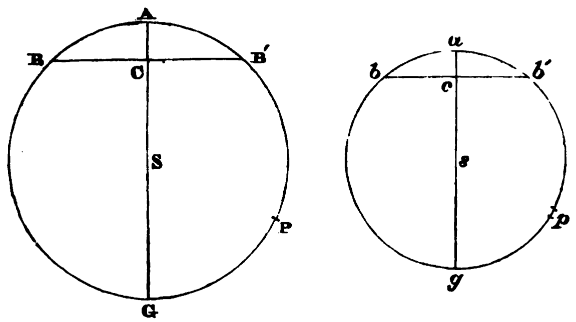
[Prop. III. is omitted, as not referring to motion about a *fixed* centre.]

Prop. IV. Theor. IV.

The centripetal forces of bodies, which describe different circles with uniform velocity, tend to the centres of the circles, and are to each other as the squares of arcs, described in the same time, divided by the radii.

Sectors of circles are proportional to the arcs on which they stand, and therefore the sectors described by the bodies are proportional to the times of describing them, since the bodies move uniformly. Hence the forces tend to the centres of the circles.

Again, let BAB', bab' be indefinitely small arcs described in equal times; join BB', bb', and draw the diameters $GSCA$, $gsca$, which bisect BB', bb' in C and c respectively. Then, (by Prop. I. Cor. 4.) ultimately,



$$\text{force at } A : \text{force at } a :: AC : ac$$
$$:: \frac{BC^2}{GC} : \frac{bc^2}{gc}$$
$$:: \frac{BB'^2}{GC} : \frac{bb'^2}{gc}$$
$$:: \frac{BAB'^2}{AG} : \frac{bab'^2}{ag};$$

but if AP, ap be any two arcs described in equal times, since the bodies move uniformly, we have

$$AP : ap = \text{ultimate value of the ratio } BAB' : bab';$$

$$\therefore \text{ force at } A : \text{ force at } a :: \frac{AP^2}{AS} : \frac{ap^2}{as};$$

$\therefore$ The centripetal forces, &c. Q.E.D.

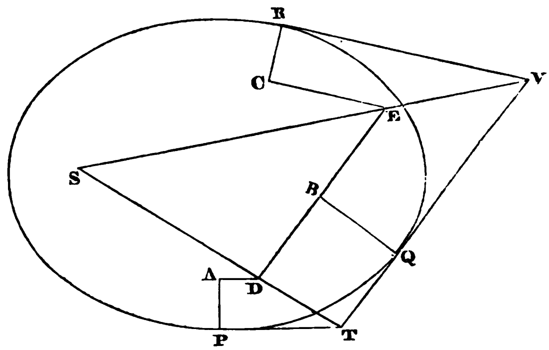
COR. 1. Since the arcs are proportional to the velocities, the forces are proportional to $\frac{(\text{velocity})^2}{\text{radius}}$; or if V be the velocity, R the radius, F the central force, then $F \propto \frac{V^2}{R}$.

COR. 2. Let P be the periodic time, then $2\pi R = VP$, (Dynamics, Art. 4. page 218.);

$$\therefore F \propto \frac{R}{P^2}.$$

PROP. V. PROB. I.

Given the velocity at all points of an orbit described by a body under the action of a central force, to find the centre.



Let the three straight lines PT, TQV, VR touch the orbit in the points P, Q, R, respectively.

Draw the lines PA, QB, RC perpendicular to these tangents, and make them inversely proportional to the velocities at P, Q, R; i.e. if V_1 V_2 V_3 are the velocities at the three points, make

$$PA : QB : RC :: \frac{1}{V_1} : \frac{1}{V_2} : \frac{1}{V_3}.$$

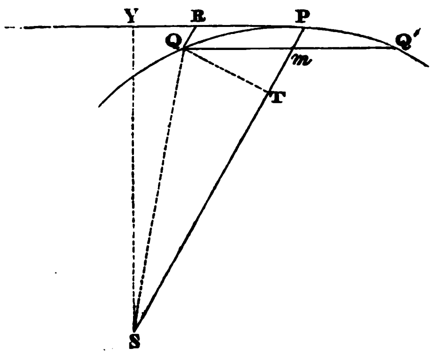
Through A, B, C draw the lines AD, DBE, CE perpendicular to PA, QB, RC. Join TD, VE: these lines produced will intersect in the centre S.

For the perpendiculars from S on the tangents PT, QT are inversely proportional to the velocities at P and Q, (Prop. I. Cor. 1.), and therefore, by construction, directly proportional to PA and QB, i. e. to the perpendiculars from D on the tangents; hence S, D, and T are in the same straight line.

Similarly, it may be shewn, that S, E, and V are in the same straight line, and therefore S is the point of intersection of TD and VE produced.

PROP. VI. THEOR. V.

If a body revolves about a fixed centre of force, and a sagitta be drawn to any very small arc, bisecting the chord of the arc and passing through the centre of force, then the force at the middle point of the arc ultimately varies directly as the sagitta and inversely as the square of the time of describing the arc.



Let QPQ' be the small arc, S the centre of force, mP the sagitta. Draw PR a tangent to the curve at P, and RQ parallel to Pm.

When the body leaves the point Q, it would if not acted upon by the central force (F) move in the direction PR; and if T be the time in which the body moves from P to Q, then the space through which it has been drawn by F is QR; hence, by Lemma X, Cor. 2,

$$F = 2 \text{ limit } \frac{QR}{T^2} = 2 \text{ limit } \frac{Pm}{T^2} \propto \text{ limit } \frac{Pm}{T^2}.$$

COR. 1. Draw QT perpendicular to SP, then will

$$F = \frac{2h^2}{SP^2} \frac{QR}{QT^2}, \text{ ultimately.}$$

For the triangle $QSP = \dfrac{SP \, . \, QT}{2}$, ultimately,

$$\therefore \; SP \,.\, QT = hT \text{ by Prop. I. Cor. 1 ;}$$

$$\therefore \; F = \frac{2h^2}{SP^2} \, \frac{QR}{QT^2}, \text{ ultimately.}$$

COR. 2. If SY be drawn perpendicular to the tangent at P, then

$$F = \frac{2h^2}{SY^2} \, \frac{QR}{PQ^2}, \text{ ultimately.}$$

For, in the limit,

$$\frac{QT}{PQ} = \frac{SY}{SP};$$

$$\therefore \; SP^2 \,.\, QT^2 = SY^2 \,.\, PQ^2,$$

$$\text{and } F = \frac{2h^2}{SY^2} \, \frac{QR}{PQ^2}, \text{ ultimately.}$$

COR. 3. If PV be the chord of curvature at P through S,

$$PV = \text{limit } \frac{PQ^2}{QR};$$

$$\therefore \; F = \frac{2h^2}{SY^2 \,.\, PV}.$$

COR. 4. If V be the velocity at P, then by Prop. I. Cor. 1.

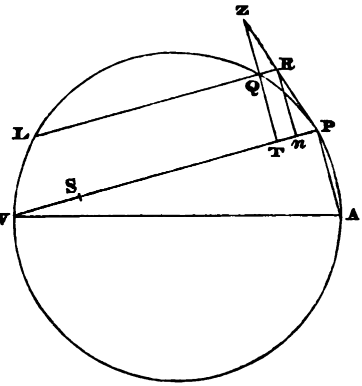
$$V = \frac{h}{SY},$$

$$\therefore \; F = \frac{2V^2}{PV}.$$

Cor. 5. Hence, if the form of the orbit in which a body moves be given, we shall be able to calculate the law of the central force. Examples of this process will be found in the following problems.

Prop. VII. Prob. II.

A body revolves in the circumference of a circle; to find the law of force tending to any given point.



Let $VQPA$ be the circle, S the given point, P the position of the body at any given time, Q a point in the orbit very near to P.

Through S draw the chord PSV, and through V the diameter VA; join PA; through Q draw ZQT perpendicular to PV, and meeting PZ the tangent at P in Z; through Q draw LQR parallel to PV, and meeting PZ in R; and lastly draw Rn parallel to QT.

$$\text{Then } F = \frac{2h^2}{SP^2} \frac{QR}{QT^2}, \text{ ultimately;}$$

but, (Euclid, III. 36.)

$$QR \cdot RL = RP^2;$$

$$\therefore F = \frac{2h^2}{RL \cdot SP^2} \frac{RP^2}{QT^2};$$

$$\text{and } \frac{RP^2}{QT^2} = \frac{RP^2}{Rn^2} = \frac{VA^2}{PV^2} \text{ by similar triangles,}$$

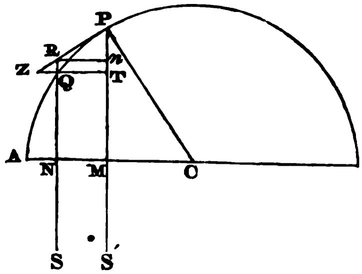
also RL ultimately $= PV$,

$$\therefore \; F = \frac{2h^2}{SP^2} \frac{VA^2}{PV^3} \propto \frac{1}{SP^2 . PV^3}.$$

COR. If the centre of force is in the circumference, $F \propto \dfrac{1}{SP^5}$.

PROP. VIII. PROB. III.

A body describes a semicircle, under the action of a force tending to a centre so distant that the force may be supposed to act in parallel lines; to find the law of force.



Let P be the position of the body at a given time, Q a contiguous position, PMS, QNS, the directions of the force at those points; draw the semi-diameter AC cutting those lines at right angles in M and N; let PRZ be the tangent at P; through Q draw ZQT perpendicular to PM and meeting PRZ in Z, and produce NQ to meet PRZ in R; join CP, and draw Rn perpendicular to PM.

Then $F \propto$ limit $\dfrac{QR}{SP^2 . QT^2} \propto$ limit $\dfrac{QR}{QT^2}$, since SP may be considered constant.

But, by Euclid, III. 36,

$$QR.(QN + RN) = RP^2,$$

$$\text{or in the limit} \quad 2\,QR . PM = RP^2,$$

$$\therefore\ F \propto \text{limit}\ \frac{RP^2}{2PM\,.\,QT^2};$$

$$\text{but}\ \frac{RP^2}{QT^2} = \frac{RP^2}{Rn^2} = \frac{CP^2}{PM^2}\ \text{by similar triangles;}$$

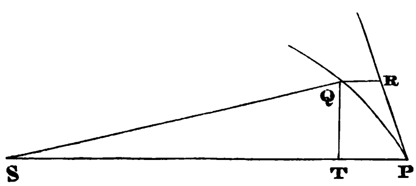
$$\therefore\ F \propto \frac{CP^2}{2PM^3} \propto \frac{1}{PM^3}.$$

Prop. IX. Prob. IV.

A body revolves in an equiangular spiral; to find the law of force tending to the centre of the spiral.

[Def. An equiangular spiral is a curve, in which the tangent at every point makes the same angle with the radius vector: *i. e.* in the figure of this proposition, SPR is a constant angle.]

Let S be the centre of force, P the position of the body at a given time, Q a point in the curve very near to P; PR the



tangent at P, QR parallel to SP, and QT perpendicular to SP.

$$\text{Then}\ F \propto \text{limit}\ \frac{QR}{SP^2.QT^2}.$$

Now, since the angle SPR is given, if we take PSQ any given small angle, we shall have all the angles in the figure $SPRQ$ given; and therefore, at whatever point of the orbit we take P, the figure $SPRQ$ will be similar, and the lines in

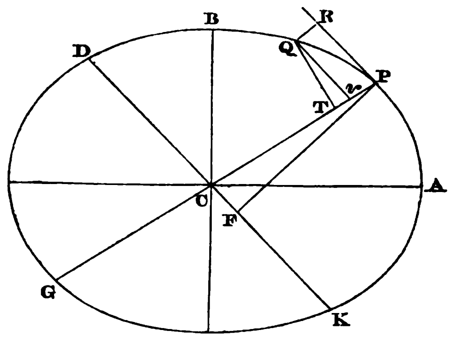
it will be proportional to any one of the homologous lines; hence QR and QT will each $\propto SP$;

$$\therefore F \propto \frac{1}{SP^3}.$$

PROP. X. PROB. V.

A body describes an ellipse; to find the law of force tending to the centre.

Let P be the position of the body at any given time, Q a point contiguous to P, PR the tangent at P, QT perpendi-



cular to the diameter PCG, DCK the diameter conjugate to PCG, PF perpendicular to DCK, Qv an ordinate to PCG.

$$\text{Then } F = \frac{2h^2}{CP^2}\frac{QR}{QT^2} \text{ ultimately.}$$

By similar triangles QTv, PFC,

$$\frac{QT}{Qv} = \frac{PF}{CP}; \therefore CP^2.QT^2 = Qv^2.PF^2;$$

$$\therefore F = 2h^2 \text{ limit } \frac{QR}{Qv^2.PF^2}.$$

But $Qv^2 = \frac{CD^2}{CP^2} Pv \cdot vG$ (Conics, Prop. VIII. page 150.)

$$= \frac{CD^2}{CP^2} QR \cdot vG;$$

$$\therefore F = 2h^2 \text{ limit } \frac{CP^2}{CD^2 \cdot PF^2 \cdot vG}$$

$$= 2h^2 \text{ limit } \frac{CP^2}{AC^2 \cdot BC^2 \cdot vG} \text{ (Conics, Prop. X. page 153.)}$$

$$= \frac{h^2}{AC^2 \cdot BC^2} \cdot CP, \text{ since } vG = 2CP \text{ ultimately,}$$

$$\text{or } F \propto CP.$$

Cor. 1. To find the periodic time in an ellipse described about a force in the centre.

Suppose $F = \mu CP$, where μ is a constant quantity depending upon the intensity of the force residing in the centre, and usually called the *absolute force* of the centre: then, by the preceding proposition, $h^2 = \mu AC^2 \cdot BC^2$.

Let P be the periodic time, that is, the time employed in describing the complete ellipse; then since the area described in a unit of time is $\frac{h}{2}$, and the whole area of the ellipse is $\pi AC \cdot BC$, we shall have

$$P = \frac{2\pi AC \cdot BC}{h} = \frac{2\pi}{\sqrt{\mu}}.$$

Hence the periodic time depends solely upon the intensity of the force in the centre.

[Cor. 2. Since P is independent of both axes of the ellipse, its value will be the same if we suppose the minor axis to be indefinitely diminished, in which case the motion will approximate to that of a body oscillating in a straight line under the action of an attractive force varying directly as the distance: hence the time of a complete oscillation of a body moving in the manner described will be $\frac{2\pi}{\sqrt{\mu}}$.]

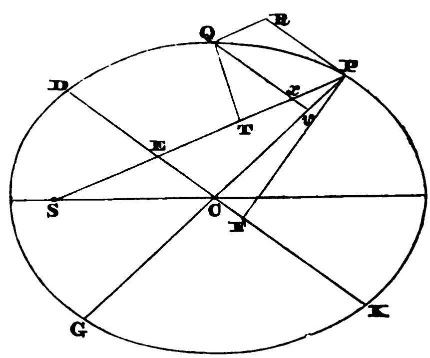
SECTION III.

ON THE MOTION OF A BODY IN A CONIC SECTION, ABOUT A CENTRE OF FORCE IN THE FOCUS.

PROP. XI. PROB. VI.

A body revolves in an ellipse; to find the law of force tending to one of the foci.

Let S be the focus of the ellipse, P the position of the body at any given time, Q a contiguous point in the orbit,



$PCGK$, DC conjugate diameters, PR the tangent at P, QR parallel to SP, Qxv to PR, QT perpendicular to SP, PF to CK, and E the point of intersection of SP and CD.

Then, $$F = \frac{2h^2}{SP^2} \frac{QR}{QT^2}, \text{ ultimately.}$$

By similar triangles QTx, PEF,

$$\frac{QT^2}{Qx^2} = \frac{PF^2}{PE^2} = \frac{PF^2}{AC^2}; \text{ (Conics, Prop. III. Cor. page 146.)}$$

and, by similar triangles Pxv, PEC,

$$\frac{Px}{Pv} = \frac{PE}{CP} = \frac{AC}{CP},$$

$$\therefore \quad QR = Px = Pv\,\frac{AC}{CP};$$

and $\dfrac{QR}{QT^2} = Pv\,\dfrac{AC}{CP}\,\dfrac{AC^2}{Qx^2 . PF^2}$

$= \dfrac{Pv}{Qv^2}\,\dfrac{AC^3}{CP . PF^2}$, ultimately,

$= \dfrac{1}{vG}\,\dfrac{CP^2}{CD^2}\,\dfrac{AC^3}{CP . PF^2}$ (Conics, Prop. VIII. p. 150.)

$= \dfrac{AC^3}{2}\,\dfrac{1}{CD^2 . PF^2}$, ultimately,

$= \dfrac{AC^3}{2\,AC^2 . BC^2}$ (Conics, Prop. X. page 153.)

$= \dfrac{AC}{2BC^2}$;

$$\therefore \quad F = \frac{h^2 AC}{BC^2} \cdot \frac{1}{SP^2} \propto \frac{1}{SP^2}.$$

[Cor. If μ be the *absolute force* of the centre,

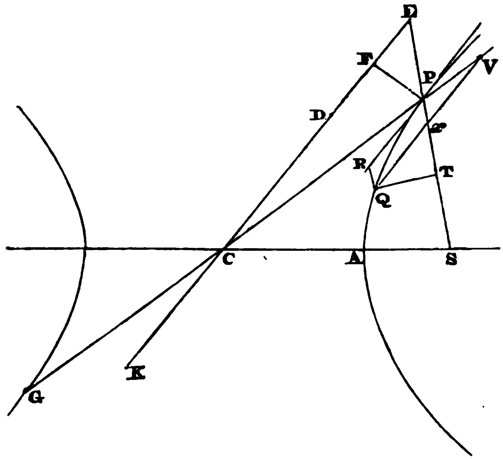
$$\therefore \quad \mu = \frac{h^2 . AC}{BC^2} = \frac{2h^2}{L},$$

where L is the latus rectum of the ellipse. (Conics, Prop. VI. Cor. page 149.)]

Prop. XII. Prob. VII.

A body moves in an hyperbola; to find the law of force tending to one of the foci.

Let S be the focus of the hyperbola, P the position of the body at any given time, Q a contiguous point in the orbit,



PCG, DCK conjugate diameters, PR the tangent at P, QR parallel to SP, Qxv to PR, QT perpendicular to SP, PF to CD produced, and E the point of intersection of SP and CD produced.

Then, $$F = \frac{2h^2}{SP^2}\,\frac{QR}{QT^2}, \text{ ultimately.}$$

By similar triangles QTx, PEF,

$$\frac{QT^2}{Qx^2} = \frac{PF^2}{PE^2} = \frac{PF^2}{AC^2} \text{ (Conics, Prop. III. Cor. page 159.)}$$

And by similar triangles Pxv, PEC,

$$\frac{Px}{Pv} = \frac{PE}{CP} = \frac{AC}{CP};$$

$$\therefore\; QR = Px = Pv\,\frac{AC}{CP};$$

and $$\frac{QR}{QT^2} = Pv\,\frac{AC}{CP}\;\frac{AC^2}{Qx^2\,.\,PF^2}.$$

$$= \frac{Pv}{Qv^2} \frac{AC^3}{CP \cdot PF^2}, \text{ ultimately},$$

$$= \frac{1}{vG} \frac{CP^2}{CD^2} \frac{AC^3}{CP \cdot PF^2} \text{ (Conics, Prop. X. page 167.)}$$

$$= \frac{AC^3}{2} \frac{1}{CD^2 \cdot PF^2}, \text{ ultimately},$$

$$= \frac{AC^3}{2AC^2 \cdot BC^2} \text{ (Conics, Prop. XI. Cor. page 169.)}$$

$$= \frac{AC}{2BC^2};$$

$$\therefore F = \frac{h^2 AC}{BC^2} \cdot \frac{1}{SP^2} \propto \frac{1}{SP^2}.$$

[Cor. As in the case of the ellipse,

$$\mu = \frac{2h^2}{L} \text{ (Conics, Prop. VI. Cor. page 162.)]}$$

Prop. XIII. Prob. VIII.

A body moves in a parabola; to find the law of force tending to the focus.

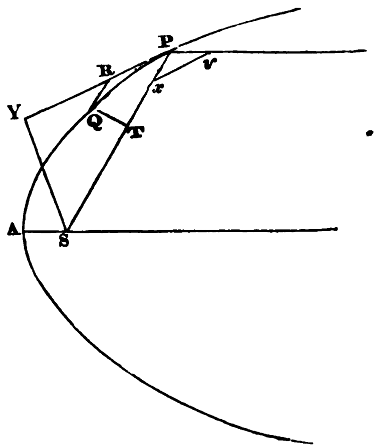
Let S be the focus of the parabola, P the position of the body at any given time, Q a contiguous point in the orbit, PRY the tangent at P, QR parallel to SP, Pxv to PR, QT perpendicular to SP, and SY to PRY.

Then $F = \dfrac{2h^2}{SP^2} \dfrac{QR}{QT^2}$, ultimately.

But $QR = Px = Pv$ (Conics, Prop. II. page 134.)

$$= \frac{Qv^2}{4SP} \quad \text{(Conics, Prop. IX. page 140.)}$$

$$= \frac{Qx^2}{4SP}, \text{ ultimately.}$$



$$\therefore \frac{QR}{QT^2} = \frac{1}{4SP} \frac{Qx^2}{QT^2} = \frac{1}{4SP} \frac{SP^2}{SY^2}, \text{ by similar triangles,}$$

$$= \frac{SP}{4AS \,.\, SP} \quad \text{(Conics, Prop. VI. Cor. page 138.)}$$

$$= \frac{1}{4AS};$$

$$\therefore F = \frac{h^2}{2AS} \cdot \frac{1}{SP^2} \propto \frac{1}{SP^2}.$$

[COR. As in the preceding cases,

$$\mu = \frac{2h^2}{L}, \quad \text{(Conics, Prop. I. page 133.)}]$$

PROP. XIV. THEOR. VI.

If any number of bodies revolve about a common centre, and the force vary inversely as the square of the distance, the latera recta of the orbits described will be as the squares of the areas described in equal times.

For we have seen in each of the three preceding propositions, that

$$\mu = \frac{2h^2}{L},$$

where μ depends upon the absolute intensity of the central force; if therefore this be given

$$L \propto h^2,$$

and since h is twice the area described in a unit of time, which may be any given time, the proposition is true. Hence if any number of bodies, &c. Q.E.D.

PROP. XV. THEOR. VII.

On the same hypothesis, the square of the periodic times in the ellipses varies as the cubes of the major axes.

Let P be the periodic time in one of the ellipses, then

$$P = \frac{2 \text{ area of the ellipse}}{h}$$

$$= \frac{2\pi AC \,.\, BC}{h},$$

$$\therefore \; P^2 = \frac{4\pi^2 AC^2 \,.\, BC^2}{\mu BC^2} AC \quad \text{(Prop. XI.)}$$

$$= \frac{4\pi^2}{\mu} AC^3 \propto AC^3.$$

Prop. XVI. Theor. VIII.

On the same hypothesis, the velocity in any of the orbits varies inversely as the perpendicular from the focus on the tangent and directly as the square root of the latus rectum.

For if v be the velocity, and p the perpendicular from the focus on the tangent, we have seen (Prop. I. Cor. 1.) that

$$vp = h,$$

$$\text{but } h \propto \sqrt{L};$$

$$\therefore\ v \propto \frac{\sqrt{L}}{p}. \quad \text{Q.E.D.}$$

Cor. 1. We have seen that $h^2 = \dfrac{\mu L}{2}$;

$$\therefore\ v^2 = \frac{\mu L}{2p^2}.$$

Now, in the parabola,

$$SY^2 = AS \,.\, SP, \text{ (Conics, Prop. VI. Cor. page 138.)}$$

$$\text{or } p^2 = \frac{L}{4} \,.\, SP;$$

$$\therefore\ v^2 = \frac{2\mu}{SP}.$$

In the ellipse,

$$SY^2 = \frac{BC^2 \,.\, SP}{2AC - SP} \text{ (Conics, Prop. IV. Cor. page 147.)}$$

$$\text{or } p^2 = \frac{L \,.\, SP}{4 - 2\dfrac{SP}{AC}};$$

$$\therefore\ v^2 = \frac{\mu}{SP}\left(2 - \frac{SP}{AC}\right).$$

Similarly, in the hyperbola,

$$v^2 = \frac{\mu}{SP}\left(2 + \frac{SP}{CA}\right). \text{ (Conics, Prop. IV. Cor. page 160.)}$$

COR. 2. Let V be the velocity with which a body would describe a circle of radius SP about the centre of force; then we may deduce the value of V from the expression for the velocity in the ellipse, given in the preceding corollary, by supposing SP and AC to be equal,

$$\therefore \; V^2 = \frac{\mu}{SP};$$

$\therefore$ in the parabola, $\quad v^2 = 2V^2$;

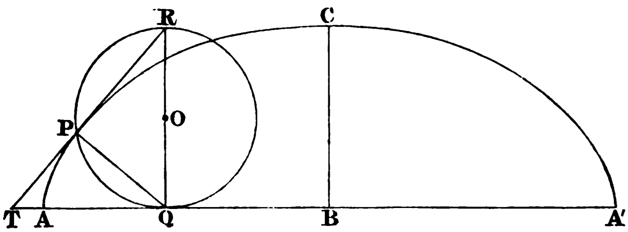
in the ellipse, $\quad v^2 = 2V^2\left(1 - \dfrac{SP}{2AC}\right)$, or $< 2V^2$;

in the hyperbola, $\quad v^2 = 2V^2\left(1 + \dfrac{SP}{2AC}\right)$, or $> 2V^2$.

APPENDIX TO SECTION II.

1. THE investigations of this section will enable us to solve the problem of finding the time of oscillation of a heavy particle, when constrained to move upon the arc of a cycloid. But before giving the solution, it will be necessary to define the cycloid, and to investigate some of its properties.

2. DEF. *A cycloid is the curve traced out by a point in the circumference of a circle, which rolls upon a given straight line.*



Thus, if a circle, of which the radius is OQ, roll on the straight line ABA', a given point P in its circumference will trace out the cycloid ACA'. It is manifest that the curve will have such a form as that exhibited in the figure; the line AA' will be equal to the circumference of the generating circle, and the curve will be symmetrical about the line BC, which bisects AA' at right angles, and which is called the axis of the cycloid.

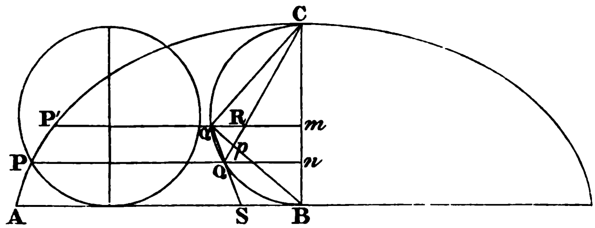
3. *To draw a tangent to a cycloid.*

Join PQ, Q being the point of the generating circle in contact with AA' at any given moment; then the generating circle moves into its next position by turning about Q, and therefore the motion of P will be for a very short space of

time the same as if it were describing a circle about Q, that is, its motion will be perpendicular to PQ.

Hence the tangent at P will be perpendicular to PQ, and will therefore pass through R the other extremity of the diameter QOR.

4. *To find the length of the arc of a cycloid.*



Let P, P' be two contiguous points in a cycloid; on the axis BC describe a semicircle, and through P, P' draw the lines Pn, $P'm$ perpendicular to BC and cutting the semicircle in Q and Q' respectively. Join CQ, CQ', BQ', $Q'Q$, and produce the last to meet AB in S; also let R, p be the intersections of CQ, $P'm$, and CQ, $Q'B$ respectively.

Then, when P' approaches indefinitely near to P, CQ will become parallel to the tangent at P, and QR will be ultimately equal to PP'. Also $Q'B$ will be ultimately perpendicular to CQ, and therefore Qp will ultimately

$$= CQ - CQ'.$$

Now SQ', ultimately $= SB$;

$$\therefore \text{ angle } SQ'B = SBQ' = mQ'B;$$

$\therefore$ in the triangles $QQ'p$, $RQ'p$, we have

angle $QQ'B = RQ'p$, and $Q'pQ = Q'pR$, (each being a right angle,) and the side $Q'p$ common;

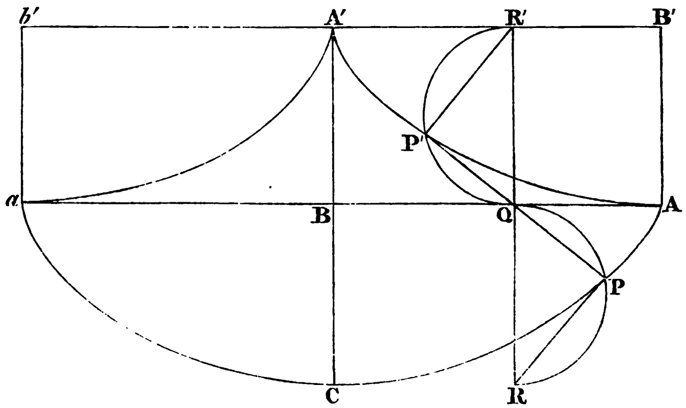
$$\therefore\ Qp = Rp,$$

$$\text{and } \therefore\ QR = 2\,Qp,$$

$$\text{or } PP', \text{ ultimately } = 2\,(CQ - CQ').$$

But PP' is the increment of the arc of the cycloi d in passing from the point P to the contiguous point P', and $CQ - CQ'$ is the corresponding increment of the chord CQ, which is equal to the chord of the generating circle touching the cycloid at P; hence *the arc of the cycloid measured from the vertex to any point equals twice the chord of the generating circle which touches the curve at that point.* In the figure of the preceding proposition $CP = 2PR$.

5. *To make a pendulum oscillate in a given cycloid.*



Let APC be a given semicycloid, having base AB and axis BC; produce CB to A', making $BA' = BC$, and complete the rectangle $A'BAB'$: with $A'B'$ as base, and AB' as axis, describe the semicycloid $A'P'A$.

Take any line $R'QR$ equal and parallel to $A'BC$, and on RQ, $R'Q$ describe the two generating semicircles QPR, $QP'R'$; join QP, PR, QP', $P'R'$.

Then the circular arc $QP = AQ$, as is manifest from the mode in which the cycloid is generated; and in like manner,

$$\text{arc } QPR = AB;$$

$$\therefore \text{ arc } PR = BQ = A'R' = \text{arc } P'R';$$

$$\therefore PR = P'R',$$

$$\text{also, } QR = QR',$$

and angle QPR = angle $QP'R'$, each being a right angle;

$\therefore$ the triangles QPR, $QP'R'$ are equal in all respects.

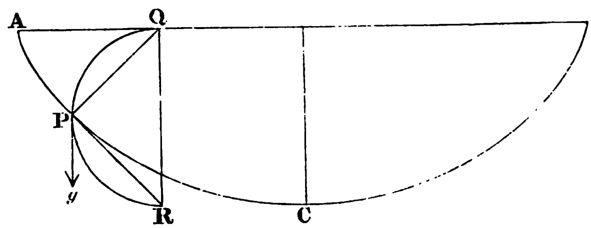
Hence angle $PQR = P'QR'$; $\therefore$ PQP' is a straight line.

Also PP' (which is a tangent to $A'P'A$ at P')

$$= 2P'Q = \text{arc } P'A.$$

Hence if a string of length $A'P'A$, fixed at A', and wrapped upon the semicycloid $A'P'A$, be unwrapped, beginning at A, a particle attached to its extremity will trace out the semicycloid APC. And by means of another semicycloid $A'a$, the particle may be made to describe the other half of the cycloid ACa.

6. *To find the time of oscillation of a heavy particle moving on the surface of a cycloid.*



Let P be the position of the particle at any time, QPR the corresponding position of the generating circle, PR the tangent at P, C the lowest point of the cycloid. Then the force, which accelerates or retards the motion of the particle, is the resolved part of the force of gravity in the direction of the tangent, that is, in the direction of PR. But gravity acts

parallel to QR, therefore the resolved part of gravity in the direction of PR

$$=g\cos PRQ=g\frac{PR}{RQ}=\frac{g}{2}\frac{PC}{QR}, \text{ (by the property of the cycloid)}$$

$$=\frac{g}{4a}\cdot PC,$$

if we call the radius of the generating circle a.

Hence the particle is always acted upon by a force tending to draw it towards C and proportional to PC, and will therefore oscillate in the same manner as a particle under the action of a central force varying directly as the distance: therefore by Newton, (Prop. X. Cor. 2),

$$\text{time of oscillation} = 2\pi\sqrt{\frac{4a}{g}} = 4\pi\sqrt{\frac{a}{g}}.$$

Cor. If we have a particle suspended by a string of length l, and made to oscillate in a cycloid by the artifice explained in the preceding proposition, then $l=4a$, and the time of oscillation

$$=2\pi\sqrt{\frac{l}{g}}.$$

It is to be observed, that by the time of oscillation is meant the time which elapses between the departure of the particle from the highest point and its return to the same.

7. When the oscillations of a pendulum are very small, we may consider the time of oscillation to be the same as if the extremities described a cycloidal arc; hence if l be the length of a pendulum we may say in general, that, provided the oscillation be small, the time

$$=2\pi\sqrt{\frac{l}{g}}.$$

The time of oscillation of a pendulum is an element which can be observed with very great accuracy, hence the observation

of a pendulum affords the best means of determining the value of the quantity g. Suppose we find by experiment the length of a pendulum which will make a semioscillation in 1″, and let L be its length, then we have

$$\pi\sqrt{\frac{L}{g}} = 1;$$

$$\therefore\ g = L\pi^2.$$

Such a pendulum is called a seconds pendulum. By this means it is ascertained that the accelerating force of gravity, though nearly the same over the earth's surface, is not accurately so. The length of the seconds pendulum, speaking without extreme accuracy, may be said to vary from the poles to the equator between the limits $39\frac{1}{5}$ and 39 inches. In the latitude of London the length is about $39\frac{1}{8}$ inches.

8. A pendulum consisting of a particle suspended by an indefinitely fine string, such as that which we have been considering, is called a *simple* pendulum. But in practice, no pendulum can be made so nearly to fulfil these conditions as to be regarded as a simple pendulum; to deduce the length of the theoretical simple pendulum from a seconds pendulum of complicated construction, requires much ingenuity, as well as the application of more complicated mathematical processes than any introduced into this work. It must suffice here to state, that the problem admits of solution to the utmost degree of accuracy.

HYDROSTATICS.

HYDROSTATICS.

1. A FLUID is a collection of material particles, which can be moved among each other by an indefinitely small force.

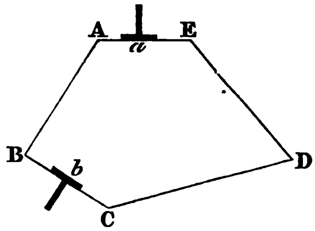
There is no fluid in nature which strictly fulfils the definition we have given; nevertheless those substances which we shall consider as fluid, fulfil it sufficiently nearly to make the conclusions founded on the definition practically correct.

2. Fluids are distinguished into *elastic* and *non-elastic*. The former class consists of those, the volume of which can be diminished by pressure, and which have an internal expansive force, in virtue of which their volume increases when not constrained by external pressure. Of this kind is *air*, and generally all *gaseous* fluids. The latter class consists of those which have not this property, and the volume of which remains the same whatever pressure they may be subjected to. Of this kind is *water*, and generally all those fluids which we term *liquid*.

These two classes of fluids are also spoken of as *compressible* and *incompressible*. Strictly speaking no known fluid is incompressible, but all ordinary liquids are sufficiently nearly so to enable us to regard them as such without sensible error.

3. As the science of Force, considered as acting on a material particle, or a system of material particles rigidly connected, divided itself into the two sciences of Statics and Dynamics, so in considering the action of force on a fluid the science will be that of *Hydrostatics* or *Hydrodynamics*, according as motion is or is not produced. The mathematical difficulty of the latter science, however, will confine us strictly to the case of a fluid at rest.

4. The characteristic property of fluids is that of *transmitting equally in all directions pressures applied at their surfaces.* Thus, suppose the figure to represent a horizontal section of a vessel containing fluid, and suppose a pressure exerted on the fluid at some part of the side AE by a piston a; then this pressure will be transmitted through the fluid, not only in one direction, as would be the case with a rigid body, but in all directions around the piston. To test the truth of this, suppose a piston b, of the same size as a, to be inserted in the side BC, then it will be found that the same force will have to be applied to the piston b, to prevent its being thrust outward, as has been applied to the piston a in order to produce the pressure on the surface of the fluid.



The same property may be proved by other experiments, so far as the nature of the case allows of experimental proof, and will be assumed as true in all that follows.

5. The pressure which a fluid exerts upon a smooth plane is necessarily perpendicular to the plane, because the pressure must be mutual, and a smooth plane is incapable of exerting any pressure parallel to its surface.

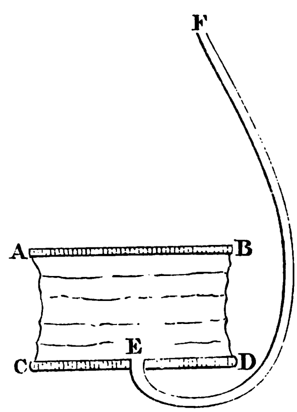
6. Having spoken of the pressure on a plane, we must explain how such pressure is measured. The pressure may be either uniform or variable, that is, it may be the same at one point as at another, or it may vary from point to point: in the former case, the pressure is measured by the pressure produced on a unit of area; in the latter, the pressure at any point is measured by the pressure which would be produced on a unit of area, if the pressure at every point of it were the same as at the proposed point. The unit of area may be any whatever, as, for instance, 1 square inch. The pressure thus measured is called *the pressure referred to a unit of surface*, and is usually denoted by the letter p. Suppose, for instance,

the pressure on each square inch of the bottom of a pail of water were the same as would be produced by putting upon it a weight of 3lbs., then $p = 3$. Also if A be the whole area pressed, and the pressure be uniform, the whole pressure will be measured by pA: thus, in the example just now taken, if the bottom of the pail contain 40 square inches, $pA = 3 \times 40 = 120$ lbs., which is the whole pressure exerted by the water.

Conversely, if the whole pressure on an area be given, and the pressure be uniform, then the pressure referred to a unit of surface will be found by dividing the whole pressure by the quantity expressing the area.

7. When we speak of the pressure at any internal point of a mass of fluid, we mean the pressure which would be exerted supposing a rigid plane were made to pass through the point in question. Or, supposing a very small portion of fluid at the given point to become rigid, then we shall have the case of a small rigid body kept in equilibrium by equal pressures on all sides of it, and the intensity of these pressures measures the pressure of the fluid at the proposed point.

8. Some very remarkable results follow from the law of equal transmission of fluid pressure, which at first, perhaps, appear somewhat paradoxical. For since, when we exert a pressure on the surface of a fluid, that pressure is transmitted equally in all directions, it is evident that the whole pressure produced on any surface will be proportional to the extent of the surface, and therefore may be increased indefinitely by increasing that surface. The following experiment exhibits this result from a very striking point of view. Suppose AB, CD to be two boards forming the ends of an air-tight leather bag, and through the lower board CD let a small tube, EF,



be introduced; then it will be found, that, by making the board AB sufficiently large, a person standing upon it and blowing into the tube, will be able to lift his own weight with ease.

ON THE EQUILIBRIUM OF NONELASTIC FLUIDS UNDER THE ACTION OF GRAVITY.

9. In the treatise on Dynamics (Art 21, page 226), we have explained what is meant by the mass of a body, and we established the formula

$$W = Mg,$$

where W is the weight of a body, M its mass, and g the accelerating force of gravity (Art. 24, page 227).

By the *density* of a body we mean the quantity of matter contained in, i.e. the mass of, a unit of its volume; so that if V be the volume of a body of uniform density ρ, and M its mass, then

$$M = \rho V,$$

$$\text{and } \therefore \; W = \rho V g.$$

It will be observed that here, as in the case of mass, (see Dynamics, Art. 21, page 226,) we are obliged to refer to the effect of gravity upon matter, and we consider two bodies of equal volume to be equally dense when their *weights* are equal, that is, when the effect of gravity upon them is the same.

The *specific gravity* of a body is the weight of a unit of its volume; so that if S be the specific gravity of a body, the volume of which is V and the weight W, then will

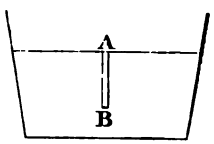
$$W = VS.$$

Comparing this with the formula last obtained, we see that

$$S = \rho g.$$

The specific gravities of different substances may be conveniently estimated with reference to some standard substance; for instance, distilled water at a given temperature; if we call the specific gravity of this standard substance 1, then those of other substances will be expressed by numbers, which give the ratios of the specific gravities of those substances to that of the standard. Thus, if the specific gravity of water is 1, that of lead is 11·35, of copper 8·9, and so of other substances.

10. *To find the pressure referred to a unit of surface at any depth below the surface of a fluid at rest.*



Let B be a point at a depth z below the surface; suppose AB to be a prism of fluid of very small transverse section a, and suppose this prism to become solid, which may evidently be done without disturbing the equilibrium; then the pressure on the base of the prism will be its weight $= \rho g a z$, if ρ be the density of the fluid. Again, let p be the pressure at B referred to a unit of surface, then the whole pressure on the base of the prism $= pa$; hence we have

$$pa = \rho g a z,$$

$$\text{or} \quad p = \rho g z.$$

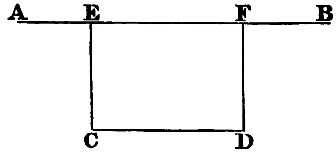
Hence the pressure at any point in the interior of a fluid at rest, is proportional to the depth below the surface.

If we suppose the surface of the fluid to be exposed to some pressure, as the pressure of the air, and we call this pressure Π, we shall have

$$p = \rho g z + \Pi.$$

11. The pressure which we have here determined is not in any definite direction, but exists in all directions around B in virtue of the fundamental property of fluids. For instance, if we have a vessel with vertical sides containing fluid, then the pressure at a given depth on the sides of the vessel will be that which we have determined, but its direction will be horizontal.

12. *The surface of a fluid at rest is a horizontal plane.*



Let AB be the surface of the fluid, CD a horizontal plane below the surface, E, F any two points in the surface, EC, FD perpendicular to CD, ρ the density of the fluid. Now, suppose a small canal of fluid joining C and D any two points in the given horizontal plane to become a solid prism; then since this prism is in equilibrium, the horizontal pressures upon its two ends must be equal, but these are the fluid pressures at C and D; hence

$$\text{fluid pressure at } C = \text{fluid pressure at } D,$$

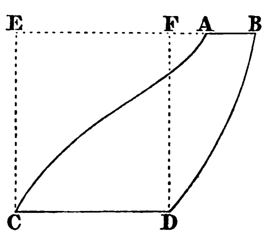
$$\text{that is, } \rho g EC = \rho g FD;$$

$$\therefore EC = FD,$$

and therefore the surface of the fluid is parallel to the horizontal plane CD, or is horizontal.

13. *To find the pressure on a plane horizontal area at any depth below the surface of a fluid at rest.*

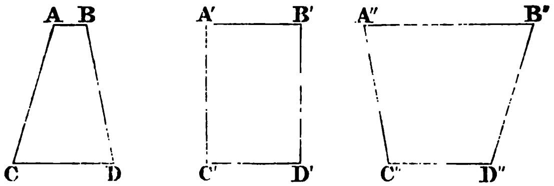
Suppose vertical lines to be drawn from all points of the circumference of the plane area to the surface of the fluid, and suppose the prism of fluid thus formed, having the given area for its base, to become solid; then the pressure on the plane area will be the same as before. But in this hypothetical case, the pressure manifestly equals the weight of the solid prism. Hence the pressure on the plane area is the weight of a column of fluid, the base of which is the area pressed, and height the depth of the area below the surface.



The proposition will be true, even when there is no such column of fluid actually superincumbent upon the plane. For, suppose we have a vessel of the shape $ACDB$, full of fluid; produce AB and draw CE, DF perpendicular to it, and suppose the part ECA to be filled with fluid; now let the side AC

of the vessel be removed, then equilibrium will still subsist and the pressure on the base CD will be the same as before. But in this case the pressure is the weight of the column $ECDF$; therefore the proposition is still true.

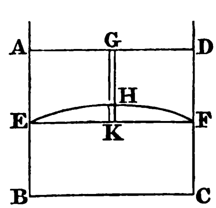
14. Hence it appears that the pressure on any plane horizontal area depends on its depth below the highest point of the fluid, and not upon the magnitude of the actual super-



incumbent mass. For instance, if we have three vessels, such as in the figure, having their bases and altitudes equal, the pressure on the bases when they are filled with fluid will be the same.

A remarkable illustration of the proposition is supplied also by this experiment; a barrel filled with water, and having a long vertical pipe of small transverse section introduced into it also filled with water, will be burst by the fluid pressure, if the pipe be of considerable length.

15. *The common surface of two fluids which do not mix is a horizontal plane.*



Let $ABCD$ be a vessel containing the fluids, and AD the horizontal surface of the upper fluid. If possible, let EHF be the common surface; draw the horizontal plane EF. Consider the equilibrium of a vertical column GHK, composed partly of one fluid and partly of the other; the pressure at K = the weight of the column GHK, but the pressure at K also equals the pressure at E which is in the same

horizontal plane with it, and therefore equals the weight of a column of fluid reaching from E to the surface.

Hence the weight of a column composed of the two fluids equals that of a column of the same height composed of only one of them; which is absurd, since the fluids are supposed to be of different densities. Therefore the common surface cannot be as we have supposed, and must be horizontal.

16. Hence, theoretically, two fluids resting the one on the other will be in equilibrium, provided their common surface is horizontal; but practically equilibrium will not subsist, unless the lower fluid be that of greater density; for if the contrary were the case, the smallest disturbance of the fluids would cause the denser fluid to descend, and the equilibrium would be destroyed: in the former case the equilibrium is said to be *stable*, in the latter *unstable*. Thus oil can rest permanently on water, but not *vice versâ*.

17. *When two fluids meet in a bent tube, the altitudes of their surfaces above the horizontal plane in which they meet, are inversely as their densities.*

For let $\rho\rho'$ be the densities, and zz' the altitudes of the fluids above the common surface; then the pressure referred to a unit of surface of the two fluids at the common surface must be equal and opposite, because there is equilibrium; call it p, then, considering the first fluid, we have (Art. 10)

$$p = \rho g z.$$

Considering the second, we have

$$p = \rho' g z';$$

$$\therefore \rho z = \rho' z',$$

$$\text{or } \frac{z}{z'} = \frac{\rho'}{\rho}.$$

18. ***The whole fluid pressure on a surface immersed in a fluid, is equal to the weight of a column of fluid, having for its base the area of the surface immersed, and for its height the depth of the centre of gravity of the surface below the surface of the fluid.***

Suppose the surface divided into a number of very small portions, each of which we may consider to be ultimately plane, and to have all its points at the same distance below the surface of the fluid. Let a_1, a_2, a_3...... be the areas of the small portions, and z_1, z_2, z_3...... their respective depths below the surface, then the pressure on a_1 is $\rho g a_1 z_1$, on a_2 the pressure is $\rho g a_2 z_2$, and so on; hence the whole pressure on the surface

$$= \rho g (a_1 z_1 + a_2 z_2 + a_3 z_3 + \ldots\ldots).$$

Let S be the area of the surface, and z the depth of its centre of gravity, then by a property of the centre of gravity*, (see Statics, Art. 44, page 209,) we shall have,

$$z = \frac{a_1 z_1 + a_2 z_2 + a_3 z_3 + \ldots\ldots}{a_1 + a_2 + a_3 + \ldots\ldots},$$

$$= \frac{a_1 z_1 + a_2 z_2 + a_3 z_3 + \ldots\ldots}{S};$$

$\therefore$ the whole pressure on the surface $= \rho g S z$,

= the weight of a column of fluid of base S and height z.

Ex. 1. An isosceles triangle is immersed in fluid, having its vertex in the surface of the fluid, and its base horizontal; find the whole pressure on the plane of the triangle.

Let the base of the triangle $= a$,

the perpendicular from the vertex on the base $= b$,

* The property referred to was proved for a number of heavy particles; when we apply the same to the centre of gravity of a surface, we must suppose it to be a physical surface of an indefinitely small thickness and having weight; it may then be considered to be made up of the component particles, the weights of which are proportional to a_1, a_2, ...

the angle at which the plane of the triangle is inclined to the horizon $= \theta$;

$$\therefore \text{ the depth of the centre of gravity } = \frac{2h}{3} \sin \theta.$$

(Statics, Art. 46, page 210.)

$$\text{Also the area of the triangle } = \frac{ab}{2};$$

$$\therefore \text{ the whole pressure } = \frac{1}{3} \rho g a b h \sin \theta.$$

Ex. 2. A cylindrical vessel, having its axis vertical, is full of fluid; find the whole pressure on the sides.

Let h be the height of the vessel, r the radius of the base; then the surface pressed $= 2\pi r h$, and the depth of the centre of gravity $= \frac{h}{2}$; $\therefore$ the whole pressure $= \pi \rho g h^2 r$.

19. The pressure on a surface, which we have been considering, is not a single pressure in a certain direction, nor does it admit, in general, of a single resultant, because the direction of the pressure on any one of the small areas into which we supposed the surface to be divided is perpendicular to that small area, and therefore varies from point to point of the surface, except in the case of a plane area. If, however, we consider only that portion of the fluid pressure which acts in any given direction, we may determine the single force in that direction, to which all the fluid pressures at different points of the surface are equivalent.

20. *When a body is immersed in a heavy fluid, the resultant of the horizontal pressures at all points of the surface of the body is zero.*

The pressure on the surface of the body will be the same in every respect as on a similar and equal portion of the fluid, supposed to be substituted for the body, and then made solid.

And this hypothetical solid will be in equilibrium under the action of its own weight, and the pressure of the fluid; but no part of its own weight acts horizontally, therefore the horizontal part of the fluid pressure must be zero.

21. *Under the same circumstances, the resultant of the vertical pressure on the body is equal to the weight, and acts through the centre of gravity, of the fluid displaced.*

Making use of the same artifice as before, the portion of fluid supposed to become solid is kept in equilibrium by its own weight and the vertical pressure of the fluid, and these must be equal and opposite forces; but the former may be supposed to act at the centre of gravity of the solidified portion, i. e. of the fluid displaced; therefore also the vertical pressure of the fluid is equal to the weight of that solidified portion, and acts through its centre of gravity.

22. *To determine the conditions of equilibrium of a floating body.*

The floating body is kept in equilibrium by its own weight acting downwards through its centre of gravity, and the pressure of the fluid acting upwards, which, as we have shewn, is equal to the weight, and acts through the centre of gravity, of the fluid displaced. Hence, when a body floats in equilibrium, the weight of the body is equal to that of the fluid displaced, and the centres of gravity of the body and of the fluid displaced are in the same vertical line.

Cor. If a body is wholly immersed in a fluid of greater specific gravity than itself, and is prevented from rising by a string or otherwise, then the force tending to raise the body is the difference between its own weight and that of the fluid displaced.

Let V be the volume of the body, S its specific gravity, S' that of the fluid; then

$$\text{pressure of the fluid upwards} = VS',$$

weight of the body acting downwards $= VS$;

$\therefore$ force tending upwards $= V(S' - S)$.

Hence we see the reason of the ascent of a balloon, when inflated with a gas specifically lighter than common air.

ON FINDING THE SPECIFIC GRAVITY OF A SUBSTANCE.

23. We have before remarked (Art. 9) that it is convenient to compare the specific gravities of substances with that of some standard substance, such as distilled water at a given temperature; we shall now explain the mode of making the comparison.

We will first consider the case of a solid body.

(1) Let the body be of greater specific gravity than the fluid.

Weigh the body first in vacuum, and then in the fluid, and let the weight in the first case be W, and in the second W'; then the weight of the fluid displaced $= W - W'$;

$$\therefore \frac{\text{specific gravity of solid}}{\text{.................. fluid}} = \frac{W}{W - W'}.$$

(2) Let the body be of less specific gravity than the fluid.

Then it must be attached to a piece of some heavy substance, the weight of which we will suppose to be w in vacuum, and w' in the fluid; and let $W_1\, W_1'$ be the weights of the two bodies attached together, in vacuum and in the fluid respectively; then

the weight of fluid displaced by the two bodies $= W_1 - W_1'$,

.... heavier $= w - w'$,

$\therefore$ .... lighter $= W_1 - W_1' - w + w'$;

$$\text{and} \therefore \frac{\text{specific gravity of solid}}{\text{.................. fluid}} = \frac{W}{W_1 - W_1' - w + w'}.$$

If the body be composed of a substance soluble in the fluid, we must inclose it in wax and proceed as before.

24. In rough experiments, founded on the preceding investigation, it will be sufficient to weigh the bodies in *air* instead of in vacuum; but in all delicate experiments, the weight of the air displaced by the body must be subtracted from its apparent weight in air.

25. *To determine the specific gravity of air.*

Let a large flask be filled with air, and weighed, and let the weight be W; again, let the air be exhausted, and the flask weighed, and its weight be W'; lastly, let the flask be filled with water, and weighed, and its weight be W''. Then the weight of air contained is $W - W'$, and of water contained $W'' - W'$,

$$\therefore \frac{\text{specific gravity of air}}{\text{.............. water}} = \frac{W - W'}{W'' - W'}.$$

26. The apparent weight of a body, resulting from an experiment made in common air, is always deceitful, except in the case of the substance weighed being of the same material as the weights used in the opposite scale of the balance.

Let V be the volume of a body weighed, S its specific gravity,

V' .... the weight, S' ....

σ the specific gravity of air.

Then we must have,

$$V(S - \sigma) = V'(S' - \sigma);$$

$$\therefore VS = V'S' \frac{1 - \dfrac{\sigma}{S'}}{1 - \dfrac{\sigma}{S}}.$$

Hence the apparent weight of a body must be multiplied by the factor $\dfrac{1-\dfrac{\sigma}{S'}}{1-\dfrac{\sigma}{S}}$ in order to get the true weight.

27. *Given volumes of substances of known specific gravities are compounded; to find the specific gravity of the compound.*

Let V V' be the volumes,

S S' the specific gravities,

σ the specific gravity of the compound.

Then, since the weight of the compound equals the sum of the weights of the constituents, we have

$$(V+V')\sigma = VS + V'S';$$

$$\therefore\ \sigma = \frac{VS+V'S'}{V+V'}.$$

28. *To compare the specific gravities of two fluids by weighing the same solid in each.*

Let W be the weight of the solid in vacuum,

W_1 its apparent weight when suspended in the first fluid,

W_2 .... second ...

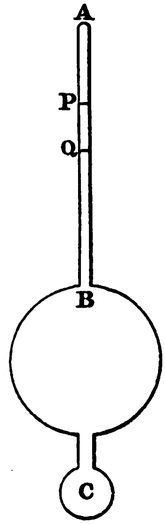
Then,

weight of the quantity of the first fluid displaced $= W - W_1$,

.... second .... $= W - W_2$;

$\therefore$ the ratio of the specific gravities $= \dfrac{W-W_1}{W-W_2}$.

29. *The specific gravities of two fluids may be conveniently compared by means of the common hydrometer.*



This instrument consists of two hollow spheres, B and C, having their centres in the axis of the graduated stem AB; the sphere C is loaded with lead, so that the instrument will float in a fluid with the stem vertical.

Let S S' be the specific gravities of two fluids which are to be compared.

V the volume of the instrument,

W its weight,

k the area of the transverse section of the stem;

and suppose that when the instrument is made to float in the two fluids, the level of the fluids in the first case is P, and in the second Q; then

$$W = S(V - k \,.\, AP),$$

$$\text{also } W = S'(V - k \,.\, AQ);$$

$$\therefore \frac{S}{S'} = \frac{V - k \,.\, AQ}{V - k \,.\, AP}.$$

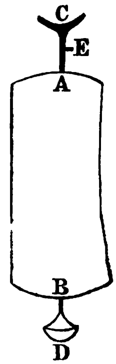
Hence by measuring AP, AQ, the ratio $\frac{S}{S'}$ is known.

30. *Nicholson's Hydrometer.*

This is a convenient instrument for comparing either the specific gravities of a solid and a fluid, or the specific gravities of two fluids.

AB is a hollow cylinder; C a dish supported by a wire AC coinciding with the axis of AB; D another dish suspended from the lower extremity of AB.

(1) To compare the specific gravities of a solid and a fluid.



Let W_1 be the weight, which placed in C causes the instrument to sink in the fluid till the surface of the fluid meets AC in a given point E. Place the solid in C, and let W_2 be the weight which must be added to make the instrument sink as deep as before. Place the solid in D, and let W_3 be the weight which must then be placed in C in order to sink the instrument to the same depth.

Then the weight of the solid $= W_1 - W_2$.

Again, the apparent weight of the solid when weighed in the fluid $= W_1 - W_3$;

$\therefore$ the weight of the fluid displaced

$$= (W_1 - W_2) - (W_1 - W_3) = W_3 - W_2;$$

$$\text{and } \therefore \frac{\text{specific gravity of solid}}{\text{.................. fluid}} = \frac{W_1 - W_2}{W_3 - W_2}.$$

(2) To compare the specific gravities of two fluids.

Let W be the weight of the hydrometer; and let W_1 W_2 be the weights which must be placed in C in order to sink the instrument down to the point E, when floating in the two fluids respectively.

The weight of the fluid displaced in the two cases will be $W + W_1$ and $W + W_2$; but the volume displaced is the same;

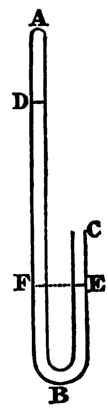
$$\therefore \text{ the ratio of the specific gravities} = \frac{W + W_1}{W + W_2}.$$

ON THE PRESSURE OF AIR AND OTHER ELASTIC FLUIDS.

31. The atmosphere or air, which surrounds the earth, produces a pressure upon all bodies immersed in it. This pressure, though very great, is not in general felt by us, be-

cause by the nature of fluid pressure it acts equally on all sides of a body submitted to it; for instance, when a man raises his hand, the downward pressure of the air above his hand is equal to the upward pressure below it, and the two therefore neutralize each other.

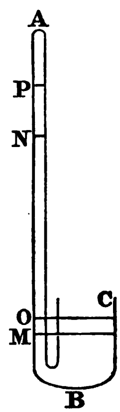
32. *To measure the pressure of the air.* (The Barometer.)



Let a bent glass tube ABC be closed at the end A, and let AB be filled with mercury. Then if the tube be placed so that AB is vertical, the mercury will descend in AB and rise in BC, leaving a vacuum above the level of the mercury in AB. Let D, E be the levels of the mercury in the two branches, and draw FE horizontal through E. Then the column of mercury FD is supported by the pressure of the air on the surface at E, and therefore if Π be the atmospheric pressure referred to a unit of surface, σ the specific gravity of mercury, and $FD = h$, we shall have

$$\Pi = h\sigma.$$

33. The barometer in common use differs slightly from the instrument just described.



The common barometer consists of a vertical closed tube AB opening into a vessel BC; a scale of inches is attached to AB. The height of the mercurial column, as shewn by such an instrument as this, is the height above a fixed horizontal plane, not above the level of the mercury in BC which is variable; hence the height will be in error, but since the area of the vessel BC is much greater than that of the tube AB the error will not be very great.

The actual error may easily be calculated, thus: let O be the zero point of graduation, and when the mercury in BC stands at that level, let N be

the level of the mercury in AB; and when the mercury in AB has risen to P, let that in BC have fallen to M, then OM is the error required. Let k K be the areas of the transverse sections of AB and BC respectively;

$$\therefore\ k \cdot PN = K \cdot OM;$$

$$\therefore\ OM = \frac{k}{K} \cdot PN = \frac{k}{K}(OP - ON);$$

and the true height of the barometer $= OP\left(1 + \frac{k}{K}\right) - \frac{k}{K}ON.$

The height of the barometer varies from about 28 to 31 inches.

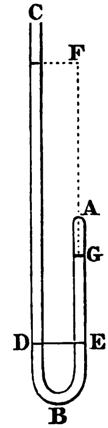
34. We remarked in the commencement of this treatise, that some fluids were elastic and some inelastic; in the latter, of which we have hitherto principally treated, the density is the same to whatever pressure the fluid may be subjected; but in elastic fluids the volume is diminished by pressure, and consequently the density increased. There will be, therefore, some relation between the volume occupied by an elastic fluid, and the pressure exerted by it, in consequence of its elasticity.

The pressure of air at a given temperature varies inversely as the space it occupies.

We shall shew how this is proved experimentally, (1) when the air is compressed for the experiment, (2) when it is rarified.

(1). Let ABC, a bent glass tube closed at A, and having its branches parallel, be placed so that the axes of the tube are vertical.

Pour mercury into the tube until it stands at the same height in the two branches, at the level DE suppose.



Now pour in more mercury, until the level in the two branches is F and G respectively.

Then if the ratio of the spaces AE, AG, occupied by the air in the two cases, be ascertained by weighing the mercury they will contain, and if h be the height of the barometer at the time of the experiment, it will be found that

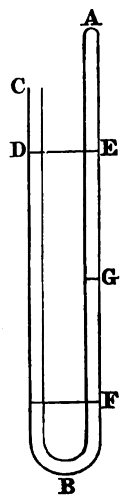
$$\frac{h + FG}{h} = \frac{\text{volume } AE}{\text{volume } AG}.$$

But if Π, Π' be the pressures of the air when occupying the spaces AE, AG respectively, and σ the specific gravity of mercury,

$$\text{then } \Pi = \sigma h,$$

$$\Pi' = \sigma(h + FG);$$

$$\therefore \frac{\Pi'}{\Pi} = \frac{\text{volume } AE}{\text{volume } AG}.$$



(2) Let a glass tube ABC, closed at A, and having the branches AB, BC parallel and nearly equal, be placed so that the axes of the branches are vertical.

Pour mercury into the tube, until the surfaces in the two branches stand at the same height DE.

Withdraw a portion of the mercury, and let the surface in the two branches then stand at F and G in the branches BC, AB respectively.

Then it is found, as in the former case, that

$$\frac{h - FG}{h} = \frac{\text{volume } AE}{\text{volume } AG}.$$

But if Π, Π' be the pressure of the air when occupying the spaces AE, AG respectively, we shall have

$$\Pi = \sigma h,$$

$$\Pi' = \sigma(h - FG);$$

$$\therefore \frac{\Pi'}{\Pi} = \frac{\text{volume } AE}{\text{volume } AG}.$$

Cor. If p be the pressure of air referred to a unit of surface when the density is ρ, we have

$$p \propto \frac{1}{\text{volume}},$$

$$\text{but } \rho \propto \frac{1}{\text{volume}};$$

$$\therefore p \propto \rho = k\rho, \text{ suppose,}$$

where k is some constant.

The same law is found to hold good in the case of all elastic fluids.

35. *If the atmosphere be supposed to be divided into indefinitely thin strata of equal thickness, the densities of the air in those strata will be in geometrical progression.*

Suppose the strata to be so thin that the density may be supposed the same throughout each; and let ρ_n, p_n be the density and pressure in the n^{th} stratum measured from the earth's surface; τ the thickness of the strata.

Then the difference of pressure in passing from the n^{th} to the $\overline{n+1}|^{\text{th}}$ stratum is the weight of a column of air of height τ;

$$\therefore p_n - p_{n+1} = \rho_n g \tau \quad \text{(Art. 10)};$$

$$\text{but } p_n = k\rho_n \quad \text{(Art. 34)},$$

$$\text{and } p_{n+1} = k\rho_{n+1};$$

$$\therefore \quad \frac{\rho_n - \rho_{n+1}}{\rho_n} = \frac{g\tau}{k};$$

in like manner, $$\frac{\rho_{n-1} - \rho_n}{\rho_{n-1}} = \frac{g\tau}{k};$$

$$\therefore \quad \frac{\rho_n - \rho_{n+1}}{\rho_n} = \frac{\rho_{n-1} - \rho_n}{\rho_{n-1}},$$

or $$\rho_{n-1}\rho_{n+1} = \rho_n^2.$$

Hence the densities $\rho_1, \rho_2, \rho_3 \dots$, and therefore also the pressures $p_1, p_2, p_3 \dots$, form a geometrical progression.

Obs. The preceding proposition is not experimentally true, for two reasons; *first*, we have considered the temperature to be the same at all heights above the earth's surface, which is not the case; and, *secondly*, we have neglected to take account of the diminution of the force of gravity as we recede from the centre of the earth. For small heights, however, the proposition may be taken as approximately true.

36. *To explain the method of finding the difference of altitude of two stations above the earth's surface by means of the barometer.*

Let x be height in feet of one station above the earth's surface,

x' .... the other ....

We may suppose the atmosphere to consist of strata of one foot thick, throughout each of which the pressure is the same, but that in passing from one to another of them the pressure diminishes in a geometrical progression. Let r be the ratio of this progression, then (making $\tau = 1$ in the last article),

$$1 - r = \frac{g}{k};$$

$$\therefore \; r = 1 - \frac{g}{k}.$$

Again, let the height of the barometer at the two stations be $h\,h'$, which will be proportional to the atmospheric pressures at the two stations;

$$\therefore\ \frac{h}{h'} = \frac{r^x}{r^{x'}} = \left(1 - \frac{g}{k}\right)^{x - x'};$$

taking logarithms, $\log \dfrac{h}{h'} = (x - x') \log\left(1 - \dfrac{g}{k}\right)$,

$$\text{or } x - x' = \frac{\log \dfrac{h}{h'}}{\log\left(1 - \dfrac{g}{k}\right)};$$

which formula, by the aid of a table of logarithms, will give us the difference of height of the two stations measured in feet*.

* The Student who is acquainted with the exponential theorem may solve this problem more completely, as follows.

Let the thickness of the strata be τ, and let $m\tau$, $n\tau$ be the heights of the two stations, and z the difference of their heights, so that

$$z = (m - n)\,\tau,$$

then $\dfrac{h}{h'} = (1 - \dfrac{g}{k}\tau)^{m-n} = (1 - \dfrac{g}{k}\tau)^{\frac{z}{\tau}}$

$$= 1 - \frac{z}{\tau}\,\frac{g}{k}\,\tau + \frac{\dfrac{z}{\tau}\left(\dfrac{z}{\tau} - 1\right)}{1\,.\,2}\left(\frac{g}{k}\tau\right)^2 - \ldots\ldots \text{ by the binomial theorem,}$$

$$= 1 - z\frac{g}{k} + \frac{z\,(z - \tau)}{1\,.\,2}\left(\frac{g}{k}\right)^2 - \ldots\ldots$$

Now make $\tau = 0$, i. e. suppose the strata to be indefinitely thin;

$$\therefore\ \frac{h}{h'} = 1 - z\frac{g}{k} + \frac{1}{1\,.\,2}\left(z\frac{g}{k}\right)^2 - \ldots\ldots$$

$$= e^{-z\frac{g}{k}} \text{ (where } e = 2{\cdot}712818\ldots\ldots) \text{ by the exponential theorem;}$$

taking logarithms,

$$\log \frac{h}{h'} = -z\frac{g}{k}\log e,$$

$$\text{or } z = \frac{k}{g}\,\frac{\log \dfrac{h'}{h}}{\log e}.$$

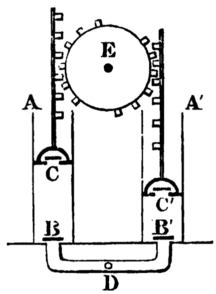
OBS. The preceding investigation explains the principle of finding heights by barometrical observations, but requires many corrections in practice to enable us to obtain accurate results. The decrease of the earth's attraction, and the change of temperature, in rising above the earth's surface give rise to the two most important corrections.

THE AIR-PUMP.

In many scientific experiments it is necessary to exhaust the air from vessels made use of. This is done by means of the air-pump; there are several varieties, some of which effect a more complete exhaustion than others, but none are capable of producing a perfect vacuum. We shall describe two constructions.

37. *Hawksbee's, or the common Air-pump.*

AB, *A′B′* are two hollow cylinders, communicating at their lower extremities by a pipe with a strong vessel or receiver, from which it is required to exhaust the air; *B*, *B′* are valves opening upwards; *C*, *C′* pistons fitted to rods which are worked by means of a toothed wheel *E*, and containing valves also opening upwards.



Suppose the piston *C* to be in its highest position, and therefore *C′* in its lowest, and suppose the density of the air in the receiver to be that of atmospheric air; then when *C* descends and *C′* rises, the valve *B* closes, and *C* opens because the pressure below becomes greater than that of atmospheric air; also *B′* opens and *C′* is closed, and the air which before occupied the receiver now occupies the receiver and the interior of the cylinder *A′ B′*, and is therefore rarified. At each stroke a similar rarification takes place; and thus the air in the receiver is gradually exhausted.

38. *To find the density of the air in the receiver after* n *turns of the wheel.*

Let A, B be the capacities of the receiver and of each of the cylinders respectively, ρ_n the density of the air after n turns, ρ the density of atmospheric air. Then after one turn the air which occupied previously the space A occupies the space $A + B$;

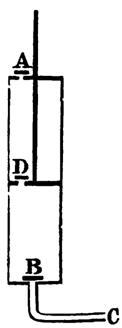
$$\therefore \quad \rho_1 (A + B) = \rho A, \text{ or } \rho_1 = \rho \frac{A}{A + B};$$

$$\text{similarly,} \quad \rho_2 (A + B) = \rho_1 A, \text{ or } \rho_2 = \rho \frac{A^2}{(A+B)^2}:$$

$$\text{and so generally,} \quad \rho_n = \rho \frac{A^n}{(A + B)^n}.$$

39. In air-pumps, which, as in the above construction, have the pistons open to the atmosphere, it is quite necessary to have two pistons; for if there were only one, the pressure of the air upon it would make it almost impossible to work the pump: by having two, as described, the pressure of the air on the pistons is in equilibrium, and the only resistance to be overcome is that arising from friction.

40. *Smeaton's Air-pump.*



AB is a hollow cylinder communicating with the receiver by a pipe *BC*; *B* is a valve opening upwards; a piston works in *AB*, having the valve *D* opening upwards; and the cylinder is closed by a plate, having a valve *A* also opening upwards.

Suppose the piston in its lowest position; when it rises the valve *B* opens, *D* shuts, and the air which occupied the cylinder is expelled through *A*; when the piston descends, *A* closes, *D* opens, *B* closes, and by raising it again the air occupying the cylinder is again expelled, and so on.

41. *To find the density of the air in the receiver after* n *ascents of the piston.*

Let A, B be the capacities of the receiver and cylinder respectively; ρ the density of atmospheric air, ρ_n the density after n ascents of the piston. Then after one ascent, the air which occupied the space A occupies the space $A + B$;

$$\therefore \quad \rho_1 (A + B) = \rho A, \text{ or } \rho_1 = \rho \frac{A}{A+B},$$

similarly,

$$\rho_2 (A + B) = \rho_1 A, \text{ or } \rho_2 = \rho \frac{A^2}{(A + B)^2},$$

and so generally,

$$\rho_n = \rho \frac{A^n}{(A + B)^n}.$$

42. In this pump only one cylinder is required, the upper surface of the piston not being exposed to the atmospheric pressure. Also the exhaustion producible is much greater than by Hawksbee's construction, because the valve D not being open to the air will open for a much longer time than the valves C, C' in the former case, which are so exposed.

43. The valves in these pumps are usually formed of a small square piece of oil silk, fastened by the four corners over an aperture in a brass plate. The receivers are glass vessels of a bell form, which stand upon a brass plate, through which the pipe enters which communicates with the cylinder or cylinders; the junction of the receiver with the brass plate is made air-tight with some greasy substance, or sometimes a disk of leather is interposed. This form of the receiver is necessary for strength, since after a few ascents of the piston the pressure of the atmosphere becomes very considerable.

ON THE THERMOMETER.

44. The thermometer is not, properly speaking, a hydrostatical instrument; nevertheless, as we have had frequently to speak of the *temperature* of fluids, it will be well to describe the instrument by means of which temperature is measured.

The effect of heat is to expand bodies under its influence; this property of bodies is taken advantage of to measure the degree of heat to which they are exposed.

45. The common thermometer consists of a glass tube, of small uniform bore, closed at one end and terminating in a bulb at the other, which together with part of the tube is filled with mercury; the part of the tube not occupied by mercury is a vacuum. A graduated scale is attached to the tube: when the thermometer is exposed to heat the mercury expands and rises in the tube; the degree of its expansion is known by the graduated scale.

46. The scale is graduated as follows. The thermometer being immersed in melting snow, a mark is made opposite to the surface of the mercury: this is the *freezing point.* The thermometer is next exposed to the steam of water boiling under a given atmospheric pressure, and a mark is made opposite to the surface of the mercury in this case: this is the *boiling point.* The interval between these two points is divided into a number of equal parts called *degrees:* in the centigrade thermometer the freezing point is called 0^0 and the boiling 100^0: in Fahrenheit's, the scale commonly used in this country, the former is marked as 32^0 and the latter 212^0.

47. *To compare the scales of two differently graduated thermometers.*

Let C^0 and F^0 denote the number of degrees indicated under the same circumstances by a centigrade and a Fahrenheit's scale. Then $F^0 - 32^0$ is the number of degrees Fahrenheit above the freezing point. Now a degree centigrade mea-

sures one hundredth part of the distance from the freezing to the boiling point, and a degree Fahrenheit measures one hundred and eightieth part of the distance.

$$\therefore \quad C \times \frac{1}{100} = (F - 32)\frac{1}{180},$$

$$\text{or } C = \frac{5}{9}(F - 32);$$

a formula by means of which we can deduce the reading of one scale from that of the other.

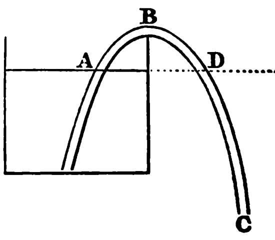
48. We can now reduce any question involving considerations of temperature to numbers; for if we speak of t degrees of temperature, we mean that the mercury in a thermometer exposed to the degree of heat in question would stand at t degrees above the zero point.

For instance, in the case of elastic fluids we have found (Art. 34. Cor.) that $p = k\rho$, provided the temperature is constant; when the temperature varies, the following is the formula given by experiment,

$$p = k\rho(1 + at),$$

where t is the temperature of the fluid, and a a small quantity, the value of which is found by experiment.

ON THE SIPHON.



49. The siphon is a bent tube ABC open at both ends. Let the tube be filled with fluid, and the shorter leg inserted into a vessel of fluid which it is required to empty, and the extremity of the other leg closed. Let the level of the surface of the fluid meet the two legs of the siphon in A and D respectively, then there will be equilibrium provided the height of B above AD is not greater than that of a column of water the weight of which is equal to the atmo-

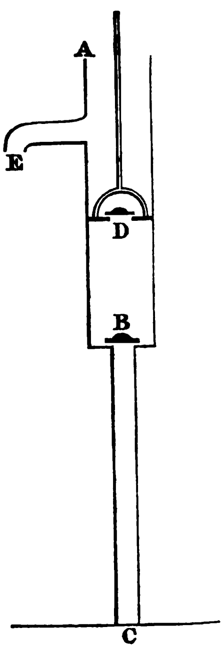
spheric pressure, and the pressure at A will be equal to that at D; consequently the pressure on the end C, which we have supposed to be closed, is greater than the atmospheric pressure, and therefore if the tube be opened the fluid will descend: the atmospheric pressure on the surface of the fluid will cause it to rise in the shorter leg, and thus a continuous stream will be produced, which will only cease when the surface of the fluid has descended to the extremity of the shorter leg of the siphon.

The limit of the height of B above the level of the surface of the fluid is about 32 feet.

ON PUMPS.

50. The pressure of the atmosphere on the surface of water is taken advantage of, for the purpose of raising it above its level. The machine by means of which this is effected, is called a *pump*; we shall describe two principal kinds.

The Common Pump.



51. AB is a cylinder, having its lower end closed with a valve B opening upwards, and connected by means of a pipe BC with the water which is to be raised. A piston, containing a valve D opening upwards, is worked in the cylinder by means of a vertical rod and a handle.

Suppose the piston to be in its lowest position; then, when it is raised, the valve B opens and a partial vacuum is produced in the cylinder and pipe, and the pressure of the atmosphere without being greater than the pressure within the pipe, the water rises, and it continues to rise until the pressure within and without become equal. When the piston descends, the valve D opens, and the air within the cylinder escapes; when it is raised, the former process is repeated, and

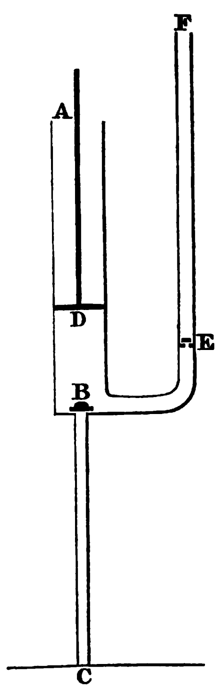
so on until the water rises to the level of the pipe E, from which it escapes.

It is obvious that the length of BC must not be greater than the height of a column of water, the weight of which is equal to the atmospheric pressure, that is, than about 32 feet.

52. The common pump is limited in respect of the height to which it can raise water; but we can raise water to any height by means of the forcing pump; it is by this means that cisterns at the higher parts of houses are supplied.

The Forcing Pump.

53. AB is a cylinder, having at its lower end a valve B opening upwards, and connected by a pipe BC with the water to be raised. From the lower part of AB a pipe EF communicates with the cistern to be filled, and this pipe is furnished with a valve E opening upwards. D is a solid piston which works in the cylinder by means of a vertical rod and handle.



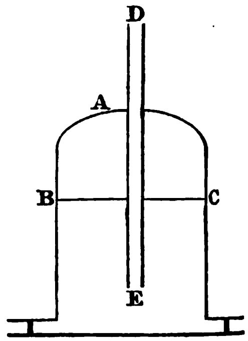
Suppose the piston to be in its lowest position; then, when it is raised, the valve B opens and the valve E is closed, and consequently a partial vacuum is produced within the cylinder and pipe; and the pressure of the atmosphere without being thus greater than that within the pipe, the water within rises, and continues to rise until the pressures within and without are equal. Let the piston be now made to descend and the process repeated, until the water has risen above the top of the pipe; then when D descends, the water in the cylinder, not being able to return on account of the valve B, is forced up the pipe EF, in which it is retained by the valve E.

This process may be continued, and by this means water may be raised to any elevation.

As in the case of the common pump, the height of the valve B above the level of the water must not exceed 32 feet.

The Fire Engine.

54. The fire engine consists of two forcing pumps, by means of which the water is forced into an air vessel ABC, from which the water can escape by the pipe ED. The air in the vessel being compressed by the water, which is forced in by the pumps, exerts a continuous pressure on the surface of the water BC, by which it is driven violently and in a continuous stream through the pipe ED. A flexible tube is attached to the mouth of the pipe ED, by means of which the stream of water can be made to play in any direction.



ON THE DIVING BELL.

55. The diving bell is a heavy chest, which is suspended by a rope, and which has its lower side open. If the bell be lowered into the water, the air within the bell will prevent the water from filling it, and consequently persons sitting on a seat inside will be enabled to breathe at considerable depths below the surface of the water.

In practice the diving bell is furnished with a flexible pipe communicating through the top of the bell with the interior, by means of which fresh air can be pumped in, and the interior thus kept as free from water as we please, while at the same time fresh air is furnished for the respiration of the divers.

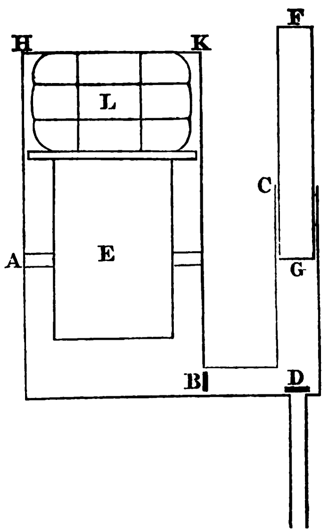
Let B be the volume of air contained by the bell, B_x the volume it contains when at the depth x below the surface, h the

height of a column of water the weight of which equals the atmospheric pressure, then (by Arts. 10 and 34)

$$\frac{B_z}{B} = \frac{h}{h+z}.$$

ON BRAMAH'S PRESS.

56. The principle of the equal transmission of fluid pressure in all directions, and the consequent possibility of increasing the total pressure on a surface to any extent by increasing the surface, supply us with the means of obtaining one of the most powerful machines in use, the application of which to the purpose of producing enormous pressure or tension is extremely valuable.



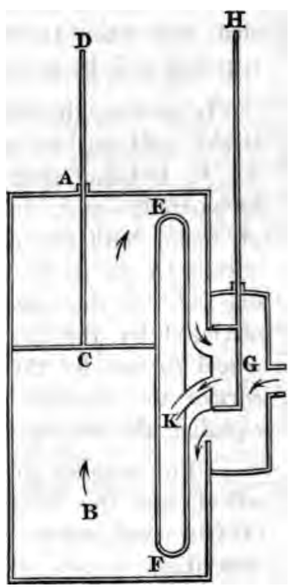
A solid cylinder E works through a watertight collar in the end of the strong hollow cylinder AB; the latter is connected by a pipe, having a valve B opening inwards, with another strong cylinder CD, which with the solid cylinder FG acting as a piston forms a forcing pump. Suppose the machine to be used for compressing a bale of goods L; then the bale is placed upon E, and is pressed by it against the very strong framework HK.

When the pump is worked the water is forced through B, which pressing on the lower surface of E causes it to rise, and to compress L; and the pressure may be increased by continuing to work the pump, the force with which the piston FG descends at each stroke being multiplied in its effect upon L by the ratio of the area of the base of the piston E to that of the piston FG.

The pressure may be immediately relieved, by allowing the water in *AB* to escape by a cock provided for the purpose.

ON THE STEAM ENGINE.

57. The account which we shall here give of the steam engine will be exceedingly brief; we shall in fact consider it principally as a hydrostatical machine, whereas the complete view of it would represent it as involving the principles of dynamics, and would require a description of a variety of ingenious contrivances which would here be out of place. It has only been by degrees that the steam engine has attained to its present perfection; we shall describe it as it was completed by James Watt, to whose genius the most important of the contrivances are due.



C is a piston working in a cylinder *AB*, the rod *CD* which communicates the motion of the piston to the machinery passing through a steam-tight collar at *A*. At *E* and *F* two pipes which communicate with the boiler enter the cylinder; suppose one of these to be open to the boiler and not the other, (in the figure the lower pipe is open to the boiler, the upper not,) then the steam rushing in through *F* below the piston will drive it up; when the piston is at the highest point of its stroke, suppose this arrangement reversed, that is, *E* to be opened to the boiler and *F* not; then the steam rushing in above the piston will drive it downward, and the steam which is below the piston will escape through *F* either into the outer air or into a vessel

provided for the purpose, it being so contrived that the arrangement which opens *E* to the boiler cuts off the communication of *F*. If this alternate opening of the pipes be effected by some contrivance, it is manifest that we shall obtain a continuous oscillating motion of the piston *C*.

The contrivances for the alternate opening of the pipes are various; in the figure we have represented a simple and common one. *G* is a steam-tight box which is made to slide up and down by means of the rod *GH*; the magnitude of this box or slide is such that when one pipe is just covered the other is just uncovered, and while one of the steam pipes, as *F* in the figure, is uncovered by it, the other (*E*) communicates through it with the vent pipe *K*. If then the rod *HG* has exactly the reverse motion of the rod *CD*, so that one shall rise when the other falls, it is evident that what was required will be done.

It is clear that the pressure on the piston throughout the stroke will not be uniform, because while one of the pipes *E*, *F*, is being shut and the other opened, they will both be partially open, and consequently there will be an influx of steam both above and below the piston; in fact, at the middle of the stroke there will be as much steam pressure on one side of the piston as on the other. This difficulty is obviated by the use of a *flywheel*, which is a large heavy wheel turned by the rod *CD*, and which by its momentum carries the machine past the half stroke, and also tends to equalize the motion throughout.

(The manner in which *CD* is connected with the flywheel, and the motion of the rod *HG* produced by an excentric crank on the axis of the flywheel, we shall not here describe.)

The steam, when allowed to escape from the cylinder, through the pipe *K*, may be permitted either to escape into the outer air, or else to flow into a closed vessel in which it is condensed, and the water formed by it pumped up again into the boiler to be reconverted into steam.

In the former case the pressure of the atmosphere is al-

lowed to act on the piston, and must be overcome by the greater pressure of the steam on the opposite side: hence, in order to produce a given effect, the elastic force of the steam, and consequently the quantity of fuel employed, must be much greater in this case, than when the steam is allowed to escape into a condenser; engines on the former construction are called *High Pressure*, on the latter *Low Pressure* engines.

The high pressure engine has also this disadvantage, that the interior of the cylinder is cooled by being open to the atmosphere, and consequently when the steam is again admitted some portion is condensed and rendered ineffective.

It is usual to cut off the steam before the piston has attained to the end of its stroke, and to allow the stroke to be completed by the elastic force of the steam already injected; by this arrangement not only is less steam required, but also the motion is more uniform.

58. An earlier and imperfect form of the steam engine was the *atmospheric* engine, in which the piston was driven up by steam, and a vacuum having been produced below it by an injection of cold water the piston was driven down by the atmospheric pressure above. The capital defect of this construction is, that the cylinder must be cooled down at every stroke, and consequently when the steam is again admitted a very large quantity is condensed, and there is an immense waste of fuel. This construction has consequently almost entirely disappeared.

OPTICS.

OPTICS.

1. The science of which we are now about to explain the principles is known as that of *common* or *geometrical* optics, in contradistinction from *physical* optics; in this latter science, we endeavour by means of a simple hypothesis concerning the constitution of light, to connect and account for the various phenomena presented to us; in the former, we are principally employed in tracing, by mathematical calculation, the results of certain experimental laws. Hence, the conclusions arrived at in the following pages, will be equally sound, whatever physical theory is adopted, and the subject will be, for the most part, one of pure geometry.

2. From a bright object light emanates in all directions, and this light we may conceive to be made up of *rays*, intending by the term *ray* to express the smallest quantity of light which can proceed in any direction; and we reason concerning rays, as though they were geometrical lines. In like manner, an object to be a source of light must be of finite, though it may be of very small dimensions; but we shall consider a bright point which is a source of light as a geometrical point.

3. Any substance which allows the transmission of light through it is called a medium. Light may proceed either through a medium or in vacuum.

4. An assemblage of rays proceeding from a luminous point is called a *pencil* of rays. The pencils which we shall consider will be conical, and the axis of the cone will be called the axis of the pencil.

A conical pencil may consist either of divergent or convergent rays; if the rays are proceeding from a luminous point, the pencil is divergent; if the rays are proceeding from

some source of light towards a point, it is convergent; if the rays are parallel, the pencil is neither divergent nor convergent.

5. When a ray of light is proceeding in a uniform medium or in vacuum, its direction is rectilinear, but when it is incident upon the surface of a medium, it is in general divided into three parts,

(1) One portion is reflected according to a regular law, and forms the *reflected ray*;

(2) Another portion enters the medium according to a regular law, and forms the *transmitted* or *refracted ray*;

(3) A third part is *scattered*, that is, is reflected in all directions without any regular law.

The first two portions mentioned are those with which we shall be hereafter concerned, the third part is that which renders the surfaces of bodies ordinarily visible.

Besides the reflected, refracted, and scattered light, there is also a certain portion *absorbed* by the medium.

In the case of polished metallic surfaces and some others, the reflected ray is the only one which sensibly exists; and, in general, the relative intensities of the reflected and refracted rays, will vary with the circumstances of the incidence, and also with the nature of the medium.

6. When a ray of light is incident upon a plane surface, the angle which its direction makes with the line perpendicular to the surface, or the *normal* to the surface, is called the *angle of incidence*, and the angles which the reflected and refracted ray respectively make with the same line are called the *angles of reflexion* and *refraction*. When a ray is incident on a curve surface, the ray will be reflected or refracted in the same manner as if it fell upon the plane which touches the surface at the point of incidence, and the angles of incidence

reflexion and refraction are those which the incident reflected and refracted ray respectively make with the normal to this plane.

7. The laws of reflexion are the following:

(1) *The incident and reflected ray lie in the same plane with the normal at the point of incidence, and on opposite sides of it.*

(2) *The angles of incidence and reflexion are equal.*

And the following are the laws of refraction:

(1) *The incident and refracted ray lie in the same plane with the normal at the point of incidence, and on opposite sides of it.*

(2) *The sine of the angle of incidence bears to the sine of the angle of refraction a ratio dependent only on the nature of the media between which the refraction takes place, and on the nature of the light.*

According to this last law, if we call the angle of incidence ϕ, and that of refraction ϕ', we shall have $\sin\phi = \mu \sin\phi$, where μ is a quantity depending upon the nature of the media and of the light; it will have for instance a certain value for refraction from vacuum into glass, another from glass into water, and so on; also it will have one value for red light, another for green, and so on. The quantity μ is called the refractive index, and is greater than 1 when refraction takes place from vacuum into a medium, and in general is greater than 1 when the refraction is from a rarer to a denser medium, and less than 1 when the opposite is the case.

8. These laws may be considered as depending for their truth upon experiment; in a treatise on Physical Optics they would be deductions from an hypothesis respecting the constitution of light, but in a treatise like the present they may be regarded as experimental truths; and we may remark that the accuracy with which observations of the heavenly bodies

can be conducted, and the perfect consistency of the results of such observations, are such as to leave no room for doubt as to the exactness of the laws on which they are founded. They are to be regarded therefore as expressing not empirical approximations, but absolute physical laws.

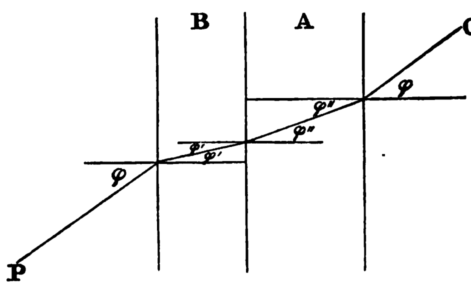
9. Prop. *If the refractive index for a medium* (A) *when light is incident upon it from vacuum be* μ, *and the index for another medium* (B) *under the same circumstances be* μ', *then when light proceeds from* (B) *into* (A) *the refractive index is* $\frac{\mu}{\mu'}$.

The proof of this proposition depends upon the two following experimental laws:

(1) If a ray of light proceed from a point P to another Q suffering any reflexions or refractions in its course, then if it be incident in the reverse direction from Q it will follow the exactly reverse course to P.

(2) If a ray pass from vacuum, through any number of media having their surfaces plane and parallel, when the ray emerges into vacuum its direction will be parallel to that which it had before incidence.

Now let ϕ be the angle of incidence from vacuum upon the medium B, ϕ' the angle of refraction, which will also be



the angle of incidence upon the medium A. Also let ϕ'' be the angle of refraction into A, which will also be the angle of

incidence upon the second bounding surface of A; and by the second of the preceding laws the angle of emergence into vacuum will be ϕ. Hence we shall have

$$\sin\phi = \mu' \sin\phi',$$

$$\text{and } \sin\phi = \mu \sin\phi'',$$

by the first of the above laws;

$$\therefore \quad \sin\phi' = \frac{\mu}{\mu'} \sin\phi'';$$

which proves the proposition.

COR. If the refractive index from vacuum into a medium be μ, that from the medium into vacuum will be $\frac{1}{\mu}$.

ON THE CRITICAL ANGLE.

10. Let ϕ be the angle of incidence of a ray within a medium, the refractive index of which is μ, and ϕ' the angle of refraction into vacuum; then

$$\sin\phi = \frac{1}{\mu} \sin\phi'.$$

From this formula if ϕ be given ϕ' may be found, and a real value will be given so long as $\sin\phi$ is less than $\frac{1}{\mu}$; but when ϕ has a value greater than that determined by the equation $\sin\phi = \frac{1}{\mu}$, the formula fails to give us a value of ϕ', it becomes in fact impossible, because the sine of an angle cannot be greater than unity. In consequence of this failure of our formula we have recourse to experiment, and we find that in reality there is no refracted ray when the angle of incidence is greater than that above assigned, the ray being wholly reflected within the medium. The angle of which the sine is $\frac{1}{\mu}$ is called the *critical angle*. The critical angle for glass is about $41^{\circ}45'$, for water about $48^{\circ}30'$.

This internal reflexion at the surfaces of media is the most complete kind of reflexion, that is to say, the reflected light is more nearly equal in intensity to the incident than in any other case.

The critical angle is sometimes called the angle of *total reflexion.*

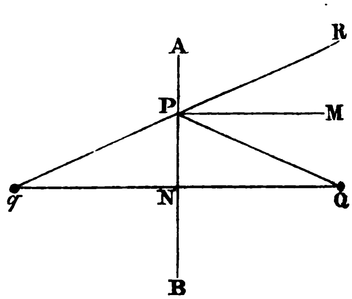
Refraction from vacuum into a medium, or from a rarer into a denser medium, is always possible.

11. We shall now proceed to investigate the effect produced upon the form of a small pencil of rays, when reflected and refracted under various conditions: the breadth of the pencil will, in general, be considered indefinitely small, for the sake of mathematical convenience, and our results must therefore be regarded as an approximation to the actual case, in which the breadth of pencils, though generally very small, is of course not indefinitely small; for many purposes the results we shall obtain will be as useful as if the approximation had not been made. We shall first consider some cases of reflexion, and then some of refraction.

ON REFLEXION AT A SINGLE SURFACE.

I. *A Plane Surface.*

12. *A conical pencil of rays is incident upon a plane reflecting surface; to determine the form of the reflected pencil.*



Let AB be a section of the surface made by a plane perpendicular to it, and passing through the luminous point Q, or focus of incidence. Draw QN perpendicular to the surface, produce it, and take $Nq = NQ$. Let QP be any incident ray, join qP and produce it to R; PR will be the reflected ray.

Draw PM perpendicular to the surface. Then in the triangles PQN, PqN, we have $NQ = Nq$, PN common, and the angles QNP, qNP equal being right angles; hence the angle $QPN = qPN = APR$: therefore also angle $QPM = RPM$, or PR is the reflected ray.

Hence the ray QP proceeds after reflexion at P, as if it came from q; and the same may be said of each other ray, therefore all the rays after reflexion proceed as if they came from q, and if the incident pencil be a cone having Q for its vertex, the reflected will also be a cone having q for its vertex.

13. We may call the point q the focus of reflexion; but it is to be observed that it is a *virtual* not a *real* focus, that is to say, the reflected rays proceed not actually, but only *as if* they came, from it. So also the line qP, which is the direction of the reflected ray PR, may be called a *virtual ray*.

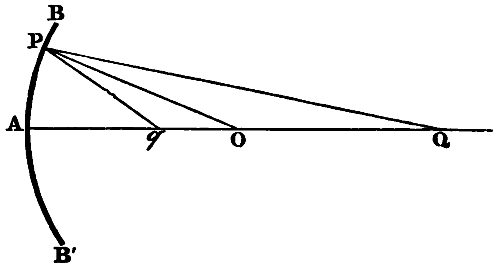
Since the rays after reflexion proceed from q as before reflexion from Q, q is sometimes called the *image* of Q.

II. *A Spherical Surface.*

14. We have seen that however large a cone of rays is incident upon a plane surface, the reflected pencil is accurately conical; but this, it is easy to see, will not be the case when rays are incident on a spherical surface. If, however, we consider the vertical angle of the incident conical pencil to be indefinitely small, the reflected pencil will also be conical, and it will be our business now to investigate the position of the vertex of the reflected cone, that of the incident being given. This vertex is called the *geometrical focus*, and if the incident rays are parallel, it is called the *principal focus of the mirror*. Also if Q be the focus of incident rays, q the geometrical focus of the reflected rays, it is manifest from the first of the two general laws quoted in Art. 9, that if q be made the focus of incident rays, Q will be the focus of reflected rays; hence Q and q are called with reference to each other *conjugate foci*.

15. ***Diverging rays are incident upon a concave spherical reflector; to find the geometrical focus.***

Let BAB' be a section of the reflector, made by a plane passing through the centre O of the sphere, and the focus



of incidence Q. Let QP be any incident ray, QOA the axis of the conical pencil: join OP, and make $OPq = OPQ$, then Pq is the reflected ray; and the ultimate position of q, when P moves up to A, will be the geometrical focus.

We have, by Euclid, VI. 3, since QPq is bisected by PO,

$$\frac{Pq}{qO} = \frac{PQ}{QO},$$

but ultimately, $Pq = Aq$, and $PQ = AQ$;

$$\therefore \quad \frac{Aq}{qO} = \frac{AQ}{QO}.$$

If we denote AQ by u, Aq by v, and AO by r, we have

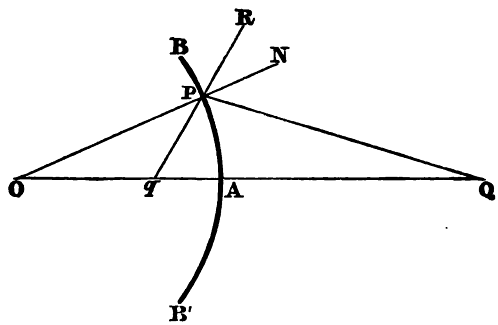
$$\frac{v}{r - v} = \frac{u}{u - r},$$

$$\text{or } \frac{r}{v} - 1 = 1 - \frac{r}{u},$$

$$\text{or } \frac{1}{v} + \frac{1}{u} = \frac{2}{r}.$$

This equation determines v when u is given.

16. *The same proposition for a convex surface.*



In this case we must draw PR, to make with OP produced to N the angle $RPN = QPN$, and produce RP to cut the axis in q.

Then since the external angle RPQ of the triangle QPq is bisected by OPN, therefore by Euclid, VI. A,

$$\frac{Pq}{qO} = \frac{PQ}{QO};$$

$$\therefore \frac{Aq}{qO} = \frac{AQ}{QO};$$

and using the same notation as before, we shall have,

$$\frac{v}{r-v} = \frac{u}{u+r},$$

$$\text{or} \quad \frac{r}{v} - 1 = 1 + \frac{r}{u},$$

$$\text{or} \quad \frac{1}{v} - \frac{1}{u} = \frac{2}{r}.$$

In this case the focus q is *virtual.*

17. In like manner we might investigate the cases of incidence of diverging rays. We shall find, however, that all four cases may be brought under one formula, by adopting

a convention respecting the sign –, as indicating direction, similar to that which we have already found so convenient in other subjects.

It will be found, that, if we suppose light to proceed from right to left across the paper, so that Q is on the right of the mirror for diverging rays, and on the left for converging, the following formulæ will result :

(1) Concave mirror, diverging rays:

$$\frac{1}{v} + \frac{1}{u} = \frac{2}{r}.$$

(2) Concave mirror, converging rays:

$$\frac{1}{v} - \frac{1}{u} = \frac{2}{r}.$$

(3) Convex mirror, diverging rays:

$$-\frac{1}{v} + \frac{1}{u} = -\frac{2}{r}.$$

(4) Convex mirror, converging rays:

$$-\frac{1}{v} - \frac{1}{u} = -\frac{2}{r}.$$

Now let us adopt this convention, that lines shall be positive or negative, according as they are measured towards the source of light, or in the opposite direction; so that u will be positive or negative, according as the incident pencil is divergent or convergent, v will be positive or negative, according as the reflected rays are divergent or convergent, and r will be positive or negative, according as the mirror is concave or convex. Then it will be seen that the four preceding formulæ will all be embraced in the following,

$$\frac{1}{v} + \frac{1}{u} = \frac{2}{r}.$$

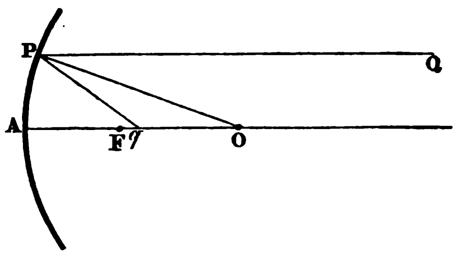
18. By making $u = \infty$ in the formula

$$\frac{1}{v} + \frac{1}{u} = \frac{2}{r},$$

we find that $v = \frac{r}{2}$, which shews that the principal focus of a mirror is half-way between the mirror and the centre of the spherical surface of which it is formed. This, however, is a sufficiently important proposition to deserve a separate investigation.

19. *To find the principal focus of a concave mirror.*

Let QP be a ray of light parallel to AO, the axis of the mirror. Join PO, O being the centre of the sphere, and



make $qPO = OPQ$, then Pq is the reflected ray corresponding to QP, and F the ultimate position of q, when P moves up to A, is the principal focus.

Then the angle $qPO = OPQ = qOP$, since Oq, PQ are parallel.

$$\therefore \quad Pq = qO;$$

and this being always true, $AF = FO$;

$$\therefore \quad AF = \frac{AO}{2}.$$

A similar demonstration is applicable to the case of a convex mirror.

AF is called the *focal length* of the mirror, and is frequently denoted by the letter f; so that $f = \frac{r}{2}$.

20. The formula $\frac{1}{v} + \frac{1}{u} = \frac{2}{r}$ may be put in a different form, by measuring the distances of Q and q from the point F.

For we have

$$QF = u - \frac{r}{2}, \quad qF = v - \frac{r}{2}.$$

$$\therefore\ u = QF + \frac{r}{2}, \text{ and } v = qF + \frac{r}{2},$$

and hence the formula $\frac{1}{v} + \frac{1}{u} = \frac{2}{r}$ becomes

$$\frac{1}{qF + \frac{r}{2}} + \frac{1}{QF + \frac{r}{2}} = \frac{2}{r},$$

$$\text{or}\quad \frac{r}{2}(QF + qF + r) = \left(qF + \frac{r}{2}\right)\left(QF + \frac{r}{2}\right)$$

$$= qF \,.\, QF + \frac{r}{2}\left\{QF + qF + \frac{r}{2}\right\};$$

$$\therefore\ qF \,.\, QF = \frac{r^2}{4} = OF^2.$$

This formula may also be proved directly in the same manner as that of Art. 15.

21. If in the formula $\frac{1}{v} + \frac{1}{u} = \frac{2}{r}$, we give u any value we obtain the corresponding value of v, which if we attend to its algebraical sign will make us acquainted with the form of the reflected pencil. Suppose for instance $r = 4$in., $u = \frac{7}{2}$in., then $\frac{1}{v} = \frac{1}{2} - \frac{2}{7} = -\frac{1}{6}$, or $v = -6$; hence the rays, which diverge upon the mirror from a point three inches and a half to the right, converge after reflexion to a point six inches to the left of the mirror. And more generally, if we suppose Q to assume all possible positions, we shall be able to find the corresponding positions of q; this we proceed to do in the following proposition.

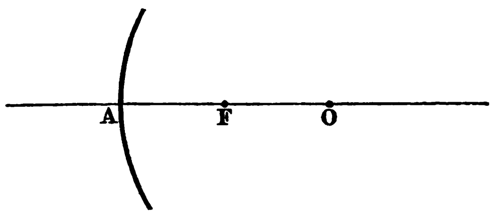
22. *To trace the corresponding positions of the conjugate foci.*

Suppose the mirror to be concave, then we have

$$\frac{1}{v} + \frac{1}{u} = \frac{2}{r}.$$

Since the sum of $\frac{1}{v}$ and $\frac{1}{u}$ is constant, as one of them increases the other must decrease, and the same will be true of their reciprocals v and u; hence Q and q always move in opposite directions.

(1) Let Q be at an infinite distance to the right of the mirror, or the incident rays parallel; then $v = \frac{r}{2}$ and q is at F.



(2) Let Q move towards O; then q moves to meet it; and at O they coincide, because when $u = r$, $v = r$.

(3) Let Q move from O towards F; then q moves to the right of O; and when Q has reached F, q is at an infinite distance, or the reflected rays are parallel, because when $u = \frac{r}{2}$, $v = \infty$.

(4) Let Q move from F towards A; then q moves from an infinite distance on the left of A to meet it; and when Q has reached A, q is there also, because when $u = 0$, $v = 0$.

(5) Let Q move to the left of A; then q moves to the right; and when Q has attained to an infinite distance, q is at F, because when $u = \infty$, $v = \frac{r}{2}$.

Q and q are now in the same positions as at first, and therefore we have traced all their corresponding positions.

The corresponding positions of the foci for a convex surface may be traced in a similar manner.

COR. It appears from the preceding investigation, that Q and q are always on the same side of F.

23. The formulæ which have been proved for a spherical mirror may be adapted to the case of a mirror formed by the revolution of any curve about its axis, by putting for the radius r the radius of curvature of the mirror at the point of incidence. Thus, suppose the mirror to be parabolical, and the latus rectum to be L, then for r we must write $\frac{L}{2}$, (see page 278), and the formula becomes

$$\frac{1}{v} + \frac{1}{u} = \frac{4}{L}.$$

The truth of the principle upon which this substitution is made will be seen at once, by considering that in the immediate neighbourhood of the point of incidence the curve and the circle of curvature may be supposed to coincide.

24. It has been observed (Art. 14) that, when a conical pencil of rays is incident on a spherical mirror, the reflected rays will not converge to or diverge from a point, unless we suppose the breadth of the pencil to be indefinitely small. There are however certain surfaces, on which if a pencil of rays of any magnitude be incident, the reflected rays will converge to or diverge from a point; such surfaces are said to be *aplanatic*.

25. *The surface formed by the revolution of a parabola about its axis, is aplanatic for rays incident parallel to its axis.*

By a property of the parabola, (Conics, Prop. II. Cor. 2, page 134), the focal distance of any point and the line drawn

through that point parallel to the axis make equal angles with the normal. Consequently, any ray incident parallel to the axis, will after reflexion pass through the focus, and therefore a pencil of rays, incident parallel to the axis of the surface formed by the revolution of the parabola about its axis, will have the focus of the parabola for the focus of reflection.

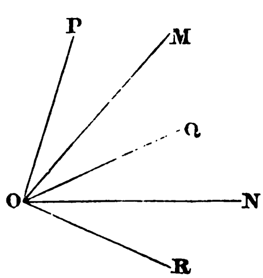
26. *The surface formed by the revolution of an ellipse about its major axis, is aplanatic for rays incident from one of its foci.*

By a property of the ellipse (Conics, Prop. II. page 144), the focal distances of any point make equal angles with the normal at that point. Consequently any ray, incident from one focus will, after reflexion, pass through the other, and therefore a pencil of rays, incident from one of the foci of the surface formed by the revolution of the ellipse about its major axis, will have the other focus for the focus of reflexion.

COMBINED REFLEXIONS.

27. *To find the deviation of a ray of light, which is reflected at the surface of two plane mirrors, inclined to each other at a given angle, in the plane perpendicular to the line of intersection of the planes.*

Through any point O draw OM, ON parallel to the normals to the two mirrors. Let PO be parallel to the incident ray; then, if we make $MOQ = MOP$, OQ is parallel to the ray after reflexion at the first surface; and again, if we make $NOR = NOQ$, OR will be parallel to the ray after reflexion at the second surface. Let α be the angle between the mirrors.

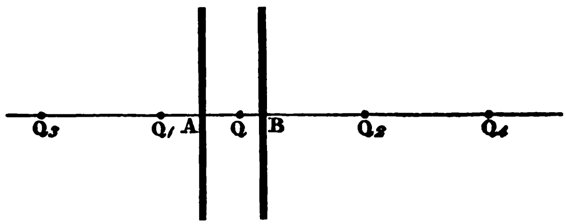


Then the deviation of the ray $= POR$

$$= POQ + QOR$$
$$= 2MOQ + 2NOQ$$
$$= 2MON$$
$$= 2a.$$

In other words, the deviation is twice the angle between the mirrors. This is an important proposition in consequence of its application to the construction of Hadley's Sextant, as shall see hereafter.

28. *A luminous point is placed between two parallel plane mirrors; to find the position of the images, which will be formed by the successive reflexions of the rays at the surfaces of the two mirrors.*



Let Q be the luminous point; through it draw AQB perpendicular to the two mirrors, and produce it both ways.

Consider the rays which fall from Q on the mirror A; if we make $AQ_1 = AQ$, an *image* of Q will be formed at Q_1 (Art. 13); again, make $BQ_2 = BQ_1$, then the rays falling as from Q_1 on the mirror B will form a second image Q_2; and if we make $AQ_3 = AQ_2$, there will be a third image at Q_3, and so on.

Let $AQ = a$, $BQ = b$: then,

$$QQ_1 = 2a,$$
$$QQ_2 = BQ + BQ_2 = BQ + BQ_1 = 2BQ + QQ_1$$
$$= 2a + 2b,$$
$$QQ_3 = AQ + AQ_3 = AQ + AQ_2 = 2AQ + QQ_2$$
$$= 4a + 2b,$$

similarly, $QQ_4 = 2BQ + QQ_3 = 4a + 4b$;

and generally, $QQ_{2n} = 2na + 2nb$,

$$QQ_{2n+1} = (2n+2)a + 2nb.$$

In like manner, if we consider the rays which fall from Q on the mirror B, we shall have a series of images, Q'_1 Q'_2...... suppose, the position of which will be determined by the formulæ

$$QQ'_{2n} = 2na + 2nb,$$

$$QQ'_{2n+1} = 2na + (2n+2)b.$$

Cor. If $a = b$, and d be the distance between the mirrors, we shall have,

$$QQ_{2n} = 2nd,$$

$$QQ_{2n+1} = (2n+1)d,$$

$$QQ'_{2n} = 2nd,$$

$$QQ'_{2n+1} = (2n+1)d;$$

which formulæ are included in this one,

$$QQ_n = nd.$$

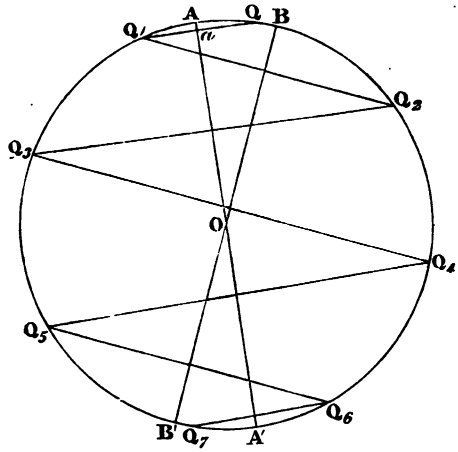
Hence the whole series of images will be equidistant from each other, and the distance between two consecutive images will be d.

29. *A luminous point is placed between two plane mirrors, inclined to each other at a given angle; to find the number and position of the images.*

Let AOA', BOB', be sections of the mirrors, made by a plane perpendicular to their line of intersection, and passing through the luminous point Q. With centre O and radius OQ describe a circle.

Consider first the rays which fall from Q on the mirror OA; draw QaQ_1 perpendicular to OA to meet the circle in Q_1, then $aQ_1 = aQ$, and therefore Q_1 is the image of Q; in like manner, if we draw Q_1Q_2 perpendicular to OB to meet the circle in Q_2, Q_2 will be the image formed by rays falling

from Q_1 on the mirror OB; and so on, the series of images Q_1 Q_2 Q_3...all lying on the circumference of the circle, which we have described.



Let the arc $AQ = \alpha$, $BQ = \beta$: then

$$QQ_1 = 2\alpha,$$

$$QQ_2 = QB + BQ_2 = QB + BQ_1 = 2\,QB + QQ_1$$

$$= 2\alpha + 2\beta,$$

similarly, $$QQ_3 = 4\alpha + 2\beta;$$

and generally, $$QQ_{2n} = 2n\alpha + 2n\beta,$$

$$QQ_{2n+1} = (2n+2)\alpha + 2n\beta.$$

In like manner there will be a series of images, Q'_1 Q'_2... suppose, formed by rays which at first fall on the mirror OB, the position of which will be determined by the formulæ,

$$QQ'_{2n} = 2n\alpha + 2n\beta,$$

$$QQ'_{2n+1} = 2n\alpha + (2n+2)\beta.$$

To determine the number of the images, we observe that reflexion will be repeated until one of the images falls at the back of the mirrors, as Q_7 in the figure; that is to say, if Q_p

be the last of the series of images $Q_1\, Q_2 \ldots, p$ must be such that AQ_p is greater than π if p be even, and BQ_p greater than π if p be odd.

Suppose p to be even and $= 2n$, then $AQ_p = a + QQ_{2n}$, and we must have

$$a + 2na + 2n\beta > \pi,$$

$$\text{or } p > \frac{\pi - a}{a + \beta}.$$

Suppose p to be odd, and $= 2n + 1$,

$$\text{then } BQ_p = \beta + QQ_{2n+1},$$

and we must have

$$\beta + (2n + 2)\, a + 2n\beta > \pi,$$

$$\text{or } p > \frac{\pi - a}{a + \beta}.$$

Hence, whether p be even or odd, it will be the whole number next greater than $\frac{\pi - a}{a + \beta}$.

Similarly, the number of images formed by rays which fall from Q on the mirror OB will be the whole number next greater than $\frac{\pi - \beta}{a + \beta}$.

Cor. If $a + \beta$, the angle between the mirrors, be an integral part of π, the number of each set of images will be $\frac{\pi}{a + \beta}$, for this will be a whole number, and $\frac{a}{a + \beta}$, $\frac{\beta}{a + \beta}$ will be proper fractions.

Suppose, for instance, $a + \beta = \frac{\pi}{3}$, then each set will consist of 3 images. If we suppose the point Q to be symmetrically situated with respect to the mirrors, the last of

the two series of images will coincide, and there will be found altogether 5 images, so that the object and images will be in the angular points of a regular hexagon.

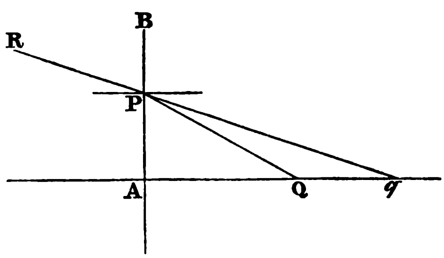
The preceding proposition contains the principle of the toy called a kaleidoscope.

ON REFRACTION AT A SINGLE SURFACE.

30. We shall now give investigations of cases of refraction, analogous to those of reflexion already considered. The figures will be drawn on the supposition of light being refracted either from vacuum into a medium, or from a rarer into a denser medium; in other words, μ will be supposed to be greater than unity, or the refracted ray to be bent *towards* the normal, unless the contrary is stated: the formulæ obtained will however be generally true.

I. *A Plane Surface.*

31. *Diverging rays are incident upon a plane refracting surface; to find the geometrical focus.*



Let Q be the focus of incidence; QA a normal to the surface AB of the refracting medium; PR the refracted ray which produced backwards meets QA in q; then the ultimate position of q, when P moves up to A, is the geometrical focus.

Let $AQ = u$, $Aq = v$, $AQP = \phi$, $AqP = \phi'$;

then, $\sin\phi = \mu \sin\phi'$,

$$\text{but } \sin\phi = \frac{AP}{PQ}, \quad \sin\phi' = \frac{AP}{Pq},$$

$$\therefore\ Pq = \mu PQ;$$

and this, being always true, will be true when P has moved up to A;

$$\therefore\ v = \mu u.$$

COR. When rays proceed from a denser into a rarer medium, μ is less than 1, and therefore q is nearer to the surface than Q. Hence when we look at an object which is under water, it will appear to be nearer to the surface of the water than it really is; for the rays which proceed from Q, when they emerge into air, proceed as if from q.

Hence also, a pole thrust partly into water will appear discontinuous at the surface of the water, and the part submerged will seem to be bent upwards; for the rays from each point of the part submerged will come to the eye, as if from a point nearer to the surface than the point itself.

32. *When a ray of light passes out of one medium into another, as the angle of incidence increases, the angle of deviation also increases.*

Let ϕ be the angle of incidence,

ϕ' .... refraction,

then the deviation $= \phi - \phi'$.

Also, let $\sin\phi = \mu \sin\phi'$;
then

$$\frac{\sin\phi + \sin\phi'}{\sin\phi - \sin\phi'} = \frac{\mu + 1}{\mu - 1},$$

$$\text{or } \frac{\sin\dfrac{\phi + \phi'}{2}\cos\dfrac{\phi - \phi'}{2}}{\cos\dfrac{\phi + \phi'}{2}\sin\dfrac{\phi - \phi'}{2}} = \frac{\mu + 1}{\mu - 1};$$

$$\therefore \tan\frac{\phi - \phi'}{2} = \frac{\mu - 1}{\mu + 1}\tan\frac{\phi + \phi'}{2}.$$

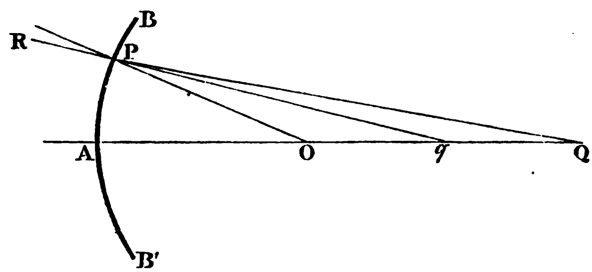
Now as ϕ increases it is manifest that ϕ' increases, because $\sin\phi' = \frac{1}{\mu}\sin\phi$; $\therefore \frac{\phi + \phi'}{2}$ $\left(\text{and } \therefore \tan\frac{\phi + \phi'}{2}\right)$ increases; and $\therefore \tan\frac{\phi - \phi'}{2}$ increases, or the deviation $\phi - \phi'$ increases.

This result is true whether μ is greater or less than unity; that is, whether the refraction takes place from a rarer into a denser medium, or from a denser into a rarer.

II. *A Spherical Surface.*

33. We shall give the investigation of the position of the geometrical focus, for the case of rays diverging upon a concave surface; the student will be able to apply the same method to all other cases, which will, however, be included in the formula which we shall obtain, by having regard to the convention respecting the negative sign explained in Art. 17.

34. *Diverging rays are incident upon a concave spherical refracting surface: to find the geometrical focus.*



Let BAB' be a section of the refracting surface, made by a plane passing through the centre O of the sphere, and the focus of incidence Q. Let QP be any incident ray, PR the corresponding refracted ray, which produced backwards

meets QOA in q. Then the ultimate position of q, when P moves up to A, will be the geometrical focus.

Join PO, and let

$$OPQ = \phi,\ OPq = \phi',\ AQ = u,\ Aq = v,\ AO = r;$$

then, in the triangle OPq,

$$\frac{Oq}{Pq} = \frac{\sin\phi'}{\sin POq},$$

and in the triangle OPQ,

$$\frac{OQ}{PQ} = \frac{\sin\phi}{\sin POQ} = \frac{\mu\sin\phi'}{\sin POq};$$

$$\therefore \frac{\mu\, Oq}{Pq} = \frac{OQ}{PQ};$$

and this, being always true, will be true when P has moved up to A, in which case

$$PQ = u,\ Pq = v,\ Oq = v - r;$$

$$\therefore \frac{\mu(v-r)}{v} = \frac{u-r}{u},$$

$$\text{or } \frac{\mu}{r} - \frac{\mu}{v} = \frac{1}{r} - \frac{1}{u},$$

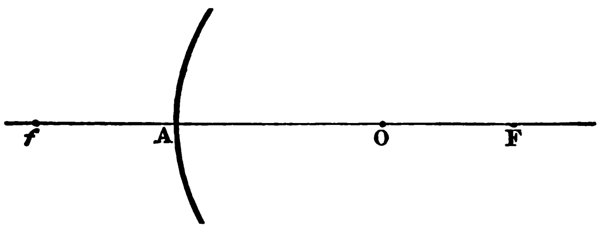
$$\text{or } \frac{\mu}{v} - \frac{1}{u} = \frac{\mu - 1}{r}.$$

OBS. It will be seen, that this formula reduces itself to that for reflexion by putting $\mu = -1$; and this is the case with all formulæ for refraction.

COR. If $u = \infty$, or the incident rays are parallel, $v = \dfrac{\mu}{\mu - 1}r$, a result which may also be easily obtained by direct investigation.

35. The relative positions of the conjugate foci are not so worthy of notice in the case of refraction at a spherical surface, as in that of reflexion, because in practice we are seldom or never concerned with refraction at a single surface: nevertheless, the student will find the following proposition worthy of attention, as tending to familiarize him with investigations of the same kind.

36. *To trace the corresponding positions of the conjugate foci.*



Suppose the surface to be concave, then we have,

$$\frac{\mu}{v} - \frac{1}{u} = \frac{\mu - 1}{r}.$$

Since the difference of $\frac{\mu}{v}$ and $\frac{1}{u}$ is constant, as one of them increases, the other must also increase, and the same will be true of v and u; hence Q and q always move in the same direction.

(1) Let Q be at an infinite distance to the right of the mirror, or the incident ray parallel; then $v = \frac{\mu}{\mu - 1} r = AF$, suppose, and q is at F.

(2) Let Q move towards O; then q also moves towards O, and at O they coincide, because when $u = r$, $v = r$.

(3) Let Q move from O towards A; then q also moves towards A, and when Q arrives at A, q arrives there also, because when $u = 0$, $v = 0$.

(4) Let Q move to the left of A; then q also moves to the left of A, and when Q has reached a point f, such that $Af = \frac{r}{\mu - 1}$, q is at an infinite distance, or the refracted rays parallel, because when $u = -\frac{r}{\mu - 1}$, $v = \infty$.

(5) Let Q move to the left of f; then q moves up from an infinite distance to the right of A, and when Q has attained to an infinite distance, q is at F, because when $u = \infty$, $v = \frac{\mu}{\mu - 1} r$.

Q and q are now in the same positions as at first, and therefore we have traced all their corresponding positions.

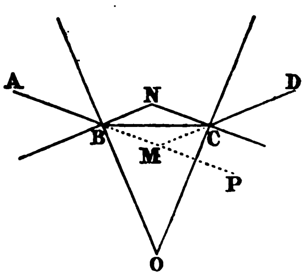
COMBINED REFRACTIONS.

37. *Refraction through a prism.*

Def. A prism is a portion of a refracting medium bounded by two planes, which terminate in a common line called the *edge* of the prism.

Def. The refracting angle of a prism is the angle of inclination of the two bounding planes.

38. ***A ray of light is refracted through a prism, in a plane perpendicular to its edge; to find the direction of the refracted ray.***



Let the plane of the paper be that in which the refraction takes place; and let $ABCD$ be the course of the ray through the prism.

Let ϕ, ϕ' be the angles of incidence and refraction at the first surface.

ψ', ψ, those at the second surface,

i the angle of the prism,

D the whole deviation.

The deviation at the first refraction is $\phi - \phi'$, and at the second $\psi - \psi'$; and, as we have drawn the figure, the two deviations evidently take place in the same directions;

$$\text{hence, } D = \phi - \phi' + \psi - \psi',$$

and we have, $\psi' + \phi' = 180^0 - BNC = i \ldots\ldots\ldots\ldots (1);$

$$\therefore \; D = \phi + \psi - i \ldots\ldots\ldots\ldots (2).$$

From the equations (1) and (2), together with the relations

$$\sin\phi = \mu \sin\phi',$$
$$\sin\psi = \mu \sin\psi',$$

$\phi' \psi \psi'$ may be eliminated, and we shall then have D expressed in terms of the given quantities ϕ and i.

If we had drawn the incident ray AB on the other side of the normal to the surface of the prism, the deviation D would have been the *difference* of the two deviations $\phi - \phi'$ and $\psi - \psi'$.

39. *To construct a prism, such that no ray shall be able to pass through it.*

Suppose the angle of the prism to be such, that a ray being incident very nearly parallel to one face, emerges very nearly parallel to the other; then it is clear, that if the refracting angle be greater than that determined by this condition, no ray will be able to pass through, because every ray which enters the prism will be incident on the second surface at an angle greater than the critical angle, and will therefore be totally reflected within the prism.

But this condition gives us, that

$$\phi = 90^0, \qquad \psi = 90^0;$$

$$\therefore \sin\phi' = \frac{1}{\mu}, \qquad \sin\psi' = \frac{1}{\mu}.$$

$$\text{Hence,} \quad i = \phi' + \psi' = 2\phi',$$

$$\text{or } \sin\frac{i}{2} = \frac{1}{\mu}.$$

The value of this angle for glass is about 83°.

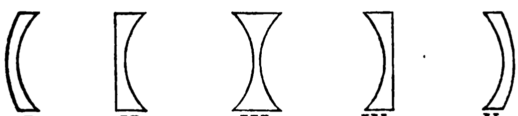
40. *Refraction through a lens.*

DEF. A lens is a portion of a refracting medium, bounded by two spherical surfaces, having a common axis.

Under this general definition, we include the case of one of the surfaces of the lens being plane, since a plane may be regarded as a sphere of infinite radius.

Lenses are of various kinds, and receive different names according to the form of the surfaces which bound them. The rule for assigning the proper designation to a lens is this: first look at the face of the lens on which the light is incident, then reverse the lens and look at the other face; combine the names, which indicate the form of the surfaces successively presented to the eye, and the combination will give the name of the lens.

If we suppose, as heretofore, that light proceeds across the paper from right to left, then the names and forms of the various lenses will be as under;



I. II. III. IV. V.

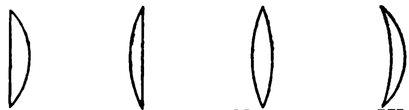
I. *Concavo-convex,*

II. *Concavo-plane,*

III. *Concavo-concave* or *double concave,*

IV. *Plano-concave,*

V. *Convexo-concave,*



VI. VII. VIII. IX.

VI. *Convexo-plane,*

VII. *Plano-convex,*

VIII. *Convexo-convex* or *double convex.*

IX. A lens such as that represented by fig. IX., according to the rule above given, would be a convexo-concave, but having very different properties from a lens of that description, it is distinguished by the name of *meniscus.*

41. *To find the geometrical focus of a small pencil of diverging rays refracted through a concavo-convex lens, the thickness of which is neglected.*

Let r, s be the radii of the surfaces of the lens.

u the distance from the lens of the focus of incidence.

V .... of the focus after refraction through the first surface.

v .... through the second.

Then, by Art. 34, we have,

$$\frac{\mu}{V} - \frac{1}{u} = \frac{\mu - 1}{r} \qquad (1).$$

Now the rays fall upon the second surface, as if they came from the geometrical focus, the position of which is determined by equation (1); but they emerge from the medium of which the lens is formed into vacuum, therefore in adapting the

formula to this case we must write $\frac{1}{\mu}$ instead of μ, (Art. 9. Cor.); hence, for the second surface,

$$\frac{1}{\mu}\frac{1}{v} - \frac{1}{V} = \frac{\frac{1}{\mu} - 1}{s},$$

$$\text{or} \quad \frac{1}{v} - \frac{\mu}{V} = -\frac{\mu - 1}{s} \qquad (2).$$

Adding (1) and (2) we have,

$$\frac{1}{v} - \frac{1}{u} = (\mu - 1)\left(\frac{1}{r} - \frac{1}{s}\right);$$

which is the formula required, and gives v when u is known.

42. The investigation is quite similar for the case of converging or parallel rays, or for any other species of lens. We have taken the particular case of rays diverging on a concavo-convex lens, because in it all lines are measured in the same direction, namely, towards the source of light; the formula for any other lens may be deduced, by attending to the convention already explained respecting the *minus* sign.

43. If we make $u = \infty$, we obtain $\frac{1}{v} = (\mu - 1)\left(\frac{1}{r} - \frac{1}{s}\right)$; the point thus determined is called the *principal focus*, or more briefly, the *focus*, of the lens. The distance of this point from the lens is called its *focal length*, and if we denote it by f, we shall have

$$\frac{1}{f} = (\mu - 1)\left(\frac{1}{r} - \frac{1}{s}\right).$$

On examining this expression it will be seen, that the first five lenses, enumerated in Art. 40, have their focal lengths positive, and the last four negative; that is to say, if parallel rays be incident upon either of the first five, they will *diverge*

after refraction, if upon either of the last four they will *converge*. Hence, lenses divide themselves naturally into two classes, which may be distinguished by the algebraical sign of f; those for which f is positive may be described in general as *concave* lenses, and those for which it is negative as *convex*. So far as the purpose of this treatise is concerned, we shall distinguish the lenses used in optical instruments only by this general character; when greater refinement is sought, there are other considerations which render the use of particular lenses in particular cases desirable.

44. It is manifest that a lens will produce the same effect, whichever face is presented to the incident light, since the formula which determines the position of the geometrical focus is not altered by supposing the face changed. Also, it appears, that a lens has two principal foci; for whichever face of the lens is exposed to parallel rays, the rays will have the same geometrical focus, and therefore a point on the axis of the lens, and at a distance f from it on either side of the lens, may be called its principal focus, and parallel rays will have one focus or the other as their geometrical focus, according as they fall on one face of the lens or the other.

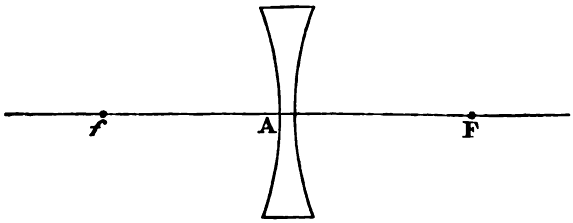
45. *To trace the corresponding positions of the conjugate foci, for a concave lens.*

The formula which determines the position of q is,

$$\frac{1}{v} - \frac{1}{u} = \frac{1}{f}.$$

Since the difference of $\frac{1}{v}$ and $\frac{1}{u}$ is constant, $\frac{1}{v}$ and $\frac{1}{u}$ must increase and decrease together, and the same will be true of v and u; hence Q and q move always in the same direction.

Let F and f be the foci of the lens, which are at a distance f from A. Then,



(1) Let Q be at an infinite distance to the right of A, or the incident rays be parallel; then q is at F, because when $u = \infty$, $v = f$, and the rays are divergent.

(2) Let Q move towards A; then q moves in the same direction, and at A they coincide, because when $u = 0$, $v = 0$.

(3) Let Q move to the left of A; then q also moves to the left of A, and when Q has reached f, q is at an infinite distance, because when $u = -f$, $v = \infty$, and the refracted rays are parallel.

(4) Let Q move to the left of f; then q moves up in the same direction from an infinite distance on the right of A, and when Q is at an infinite distance, q is at F, because when $u = \infty$, $v = f$.

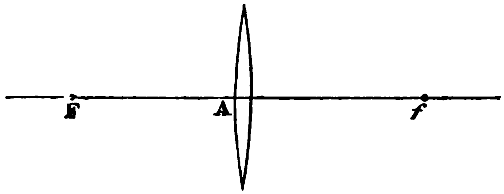
Q and q are now in the same positions as at first, and we have therefore traced all their corresponding positions.

In the figure we have represented a double concave lens, but the same investigation is applicable to all lenses for which f is positive. On account of the importance to the student of perfect familiarity with the relative positions of the conjugate foci, we shall consider at length the case of a convex lens, or in general of a lens for which f is negative.

46. *To trace the corresponding positions of the conjugate foci, for a convex lens.*

The formula which determines the position of q is,

$$\frac{1}{v} - \frac{1}{u} = -\frac{1}{f}.$$



Q and q move in the same direction, as before. Let F and f be the foci of the lens, which are at a distance f from A. Then,

(1) Let Q be at an infinite distance to the right of A, or the incident rays parallel; then q is at F, because when $u = \infty$, $v = -f$, and the refracted rays are convergent.

(2) Let Q move towards f; then q moves to the left of F, and when Q has reached f, q is at an infinite distance, because when $u = f$, $v = \infty$, and the refracted rays are parallel.

(3) Let Q move from f towards A; then q moves up from an infinite distance on the right of the lens towards A, and at A they coincide, because when $u = 0$, $v = 0$.

(4) Let Q move to the left of A; then q also moves to the left of A, and when Q is at an infinite distance, q is at F, because when $u = \infty$, $v = -f$.

Q and q are now in the same positions as at first, and therefore we have traced all their corresponding positions.

47. In examining the preceding investigations, it will be seen, that the concave lens always diminishes the convergency or increases the divergency of rays, and that the convex lens has exactly the reverse effect. This very important property of lenses admits however of simple direct proof, as in the following proposition.

48. *A concave lens diminishes the convergency or increases the divergency of a pencil of rays, and a convex lens produces the opposite effect.*

For the concave lens,

$$\frac{1}{v} = \frac{1}{u} + \frac{1}{f}.$$

Suppose u positive, or the incident rays divergent, then $\frac{1}{v}$ is greater than $\frac{1}{u}$, or v less than u; therefore the geometrical focus is nearer to the lens than the focus of incidence, or the refracted rays are more divergent. If u is negative, or the incident rays convergent, v is either positive or a greater negative quantity than u; therefore the refracted rays are either divergent, or less convergent than the incident.

For the convex lens,

$$\frac{1}{v} = \frac{1}{u} - \frac{1}{f}.$$

From the discussion of which equation, the opposite conclusions to the above are deduced. Hence the proposition is true.

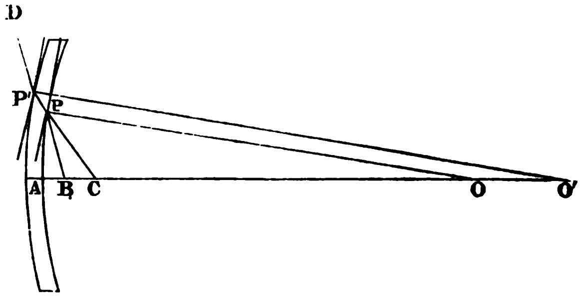
49. When reflexion takes place at a spherical surface, it is evident that a ray which passes through the centre of the sphere suffers no deviation, because it falls upon the surface in a direction coinciding with the normal at that point: there is a similar point on the axis of a lens, through which, if the direction of a ray during its course through the lens pass, it will emerge parallel to the direction of incidence, and therefore if we neglect the thickness of the lens will suffer no deviation: this point by analogy is called the *centre of the lens*.

50. *To find the centre of a lens.*

Let the lens be concavo-convex. OO' the centres of the first and second surfaces respectively.

Draw any line OP to the first surface, and $O'P'$, parallel to OP, to the second surface; join PP'; and let $BPP'D$ be

the course of a ray through the lens; produce PP' to meet the axis in C.



Then, since OP, $O'P'$ are parallel, the planes perpendicular to them are parallel, and therefore the ray $BPP'D$ passes as through a medium bounded by parallel surfaces, and $P'D$ is parallel to BP.

Again, by similar triangles, CPO, $CP'O'$,

$$\frac{CO}{PO} = \frac{CO'}{P'O'},$$

or, $\dfrac{r - AC}{r} = \dfrac{s - t - AC}{s}$, where t is the thickness of the lens;

$$\therefore AC = \frac{rt}{s - r}.$$

This formula determines the position of C, the centre of the lens, which as we see depends on the form of the lens only.

51. If the surfaces of the lenses have their convexities turned in opposite directions, that is, if r and s have different algebraical signs, the distance AC is very small when t is very small; and hence, in the case of double convex or double concave lenses, which are very thin, we may consider in practice that the centre coincides with the middle point A of the lens, and therefore that a ray passing through A suffers no deviation.

But the centre of a concavo-convex or of a meniscus lens may be at a considerable distance from the lens, if r and s be nearly equal.

52. Sometimes two lenses are placed on the same axis, and very near to each other, so as to serve the purpose of one lens: the focal length of the lens, which would produce the same effect on a pencil of rays as the two together, is called the *focal length of the combination*.

53. *To find the focal length of a combination of two lenses, omitting their thickness.*

Let ff' be the focal lengths of the lenses, F that of the combination.

Suppose rays to diverge from a point at distance u from the first lens, and let V be the distance of the geometrical focus; from this focus rays diverge upon the second lens, let v be the distance of the geometrical focus after this second refraction.

Then we have,

$$\frac{1}{V} - \frac{1}{u} = \frac{1}{f},$$

$$\text{and } \frac{1}{v} - \frac{1}{V} = \frac{1}{f'};$$

adding these equations,

$$\frac{1}{v} - \frac{1}{u} = \frac{1}{f} + \frac{1}{f'}.$$

But if the rays diverged upon a lens of focal length F, the formula would be

$$\frac{1}{v} - \frac{1}{u} = \frac{1}{F}.$$

Hence, in order that this lens may be equivalent to the combination, we must have

$$\frac{1}{F}=\frac{1}{f}+\frac{1}{f'}.$$

If the first lens be concave, and the second convex, we should have in like manner,

$$\frac{1}{F}=\frac{1}{f}-\frac{1}{f'};$$

and so in other cases.

A similar investigation is applicable to three or more lenses.

ON IMAGES FORMED BY REFLEXION OR REFRACTION.

54. When light is incident from a luminous point upon a reflecting or refracting surface, or upon a combination of surfaces, whether plane or spherical, the investigations of the preceding pages enable us to determine the focus of the reflected or refracted rays; and the focus so determined is to be considered as the *image* of the luminous point, that is, the rays will proceed from it as from a luminous point, and will therefore produce the same effect upon our organs of vision as an actual luminous point. Our investigations only apply strictly to the case of an indefinitely small conical pencil of rays, incident *directly* on the reflecting or refracting surface, that is, having its axis coincident with the axis of the surface; if the incidence be *oblique*, that is, if the axis of the pencil be inclined at some angle to the axis of the surface, the reflected or refracted pencil will manifestly be altered in its form, and our formulæ will not be strictly correct. Nevertheless, since in practice the obliquity is generally small, and since the consideration of the general question of the form of oblique pencils would lead us into calculations

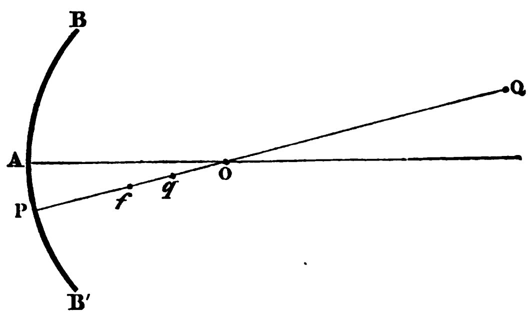
more complicated than we desire to introduce into this treatise, we shall suppose that the formulæ already established for direct, will also hold for oblique incidence, a supposition which the student will bear in mind is only approximately true.

55. Having then solved the problem of finding the image of a point, we may now proceed to consider the more general one of finding the image of any object formed by reflexion or refraction. We may conceive any object to be made up of physical points, each of which is a focus of incidence, and has a corresponding focus of reflexion or refraction; if then we construct the geometrical foci, corresponding to all different points of the object, we shall have the image of the object required.

56. We shall confine our attention almost exclusively to the case of the image of a small straight line, because this is the simplest figure which the object can have, and the determination of the image in this case will be sufficient for our purpose, when we come to apply our results to the construction of optical instruments.

57. We shall first, however, shew how to apply our formulæ to the finding of the image of a point, from which the rays fall obliquely on a mirror or a lens.

Let BAB' be a concave spherical mirror, and AO its axis. Let Q be a luminous point not on the axis of the mirror.



Draw QOP through the centre O of the mirror, and take q such that

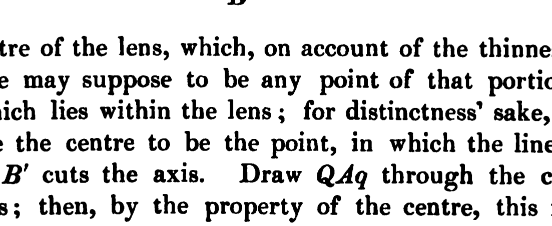
$$\frac{1}{Pq} + \frac{1}{PQ} = \frac{2}{PO};$$

then q will be the image of Q.

The position of q may be determined very easily and sufficiently nearly for many purposes, by means of the investigation of Art. 22. Bisect OP in f, then if Q were at an infinite distance q would be at f; but as Q moves towards O, q also moves towards O, therefore when Q has the position given to it in the figure, q is somewhere between f and O. In like manner, if Q were between O and f, q would be on PO produced; and if Q were between f and P, q would be on OP produced.

In like manner we may find the image of a luminous point in the case of a convex mirror.

58. Next let us consider a lens. Let BAB' be a double convex lens; Q a luminous point not on its axis. Let A be



the centre of the lens, which, on account of the thinness of the lens, we may suppose to be any point of that portion of its axis which lies within the lens; for distinctness' sake, we will suppose the centre to be the point, in which the line joining B and B' cuts the axis. Draw QAq through the centre of the lens; then, by the property of the centre, this ray will undergo no deviation, and consequently the image of Q must be somewhere on the line QAq. If we suppose the formula proved in Art. 41, to apply to this case of oblique incidence,

the distance (Aq) of the image from A will be given by the formula,

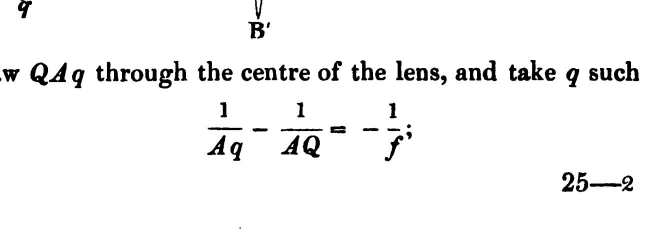
$$\frac{1}{Aq} - \frac{1}{AQ} = -\frac{1}{f}.$$

But, as in the case of reflexion, we may determine the position of q sufficiently nearly for many purposes, by considering that if $Af = f$, (the focal length of the lens,) q is at an infinite distance from A, and that as Q moves away from A, q moves from the left towards A, and will have some position to the left of F.

In like manner, we may determine the image of a luminous point, in the case of a double concave or any other lens.

59. The same method may be adapted to the case, in which a small pencil of rays falls *excentrically* on a lens, that is, in which the axis of the incident pencil does not pass through the centre of the lens. For we may conceive a complete pencil, having its axis passing through the centre, to fall upon the lens, and we may determine the image in this case, and the same will be the position of the image when the pencil is so restricted that it becomes excentrical; the difference will be, that there will be fewer rays diverging from, or converging to, the geometrical focus.

Suppose, for instance, a very small excentrical pencil falls from Q on the portion aa' of a double convex lens BAB':



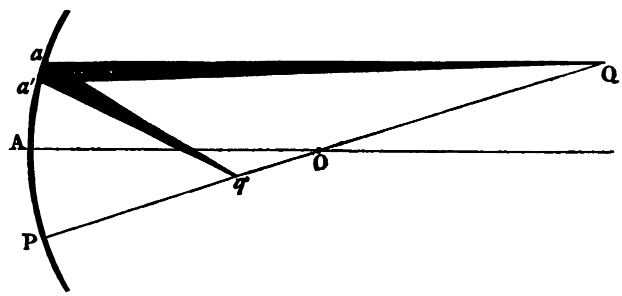
draw QAq through the centre of the lens, and take q such that

$$\frac{1}{Aq} - \frac{1}{AQ} = -\frac{1}{f};$$

then the rays incident from Q will, after refraction through the lens, converge to q. If $AQ = f$, $Aq = \infty$, and the refracted rays are parallel; hence, when a small pencil of rays is incident on a convex lens from a point at a distance from it equal to its focal length, we must draw a line through the point and the centre of the lens, and the rays will emerge parallel to that straight line.

60. A similar method is applicable to a small excentrical pencil, incident on a mirror. A pencil is in this case said to be excentrical, when the point in which its axis meets the mirror is not that in which the axis of the mirror meets it.

Let Q be the origin of a small pencil, incident on the small portion aa' of the mirror. Draw $QOqP$ through the



centre O of the spherical surface, and take the point q such that

$$\frac{1}{Pq} + \frac{1}{PQ} = \frac{2}{r};$$

then the rays of the small excentrical pencil will, after reflexion, converge to q.

61. We are now prepared to consider the formation of the image of a straight line, placed before a reflecting surface, or before a lens.

The image of an object, placed before a plane reflector, will manifestly be precisely similar to the object, and each point of the image will be as far behind the mirror, as the

point of which it is the image is distant from the mirror. This follows at once from Art. 12.

When a small straight line is placed before a spherical mirror, its image, that is, the locus of the geometrical foci corresponding to the different points of it, will evidently not be a straight line. If the line be indefinitely distant from the mirror, the image will evidently be a small arc of a circle; in other cases it may be shewn to be a portion of a conic section, but this we shall not do, since we shall not be concerned with the particular form which the image assumes: if however the object be supposed to be extremely small, it will be sufficient to consider its image to be also a straight line, and the only point with which we shall engage ourselves will be the determination of the position of this straight line.

62. We shall here give results, which the student will have no difficulty in verifying for himself.

Concave Mirror. When the object is at a distance from the mirror greater than its radius, the image will be small, inverted, and between the centre and principal focus. When the object is between the centre and principal focus, the image will be magnified, inverted, and at a distance from the mirror greater than its radius. When the object is between the principal focus and the mirror, the image will be magnified, erect, and behind the mirror.

Convex Mirror. The image will always be small, erect, behind the mirror, and between the mirror and its principal focus.

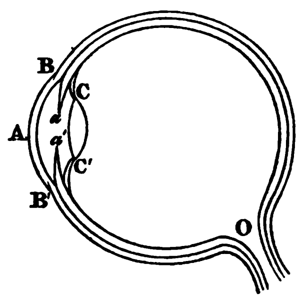
Concave Lens. The image will be small, erect, on the same side of the lens as the object, and between the lens and its principal focus.

Convex Lens. When the object is at a distance from the lens greater than its focal length, the image is on the opposite side of the lens, inverted, and at a distance from the lens greater than its focal length. When the object is between the lens and its principal focus, the image is erect, and on the same side of the lens as the object.

ON THE EYE.

63. The theory of the formation of images, which we have been explaining, is applicable to the explanation of the construction of the eye; which, to take the simplest view, may be conceived of as a convex lens, by means of which, images of external objects are formed upon a screen behind it, these images affecting the brain by means of nerves, and in some inexplicable manner conveying to the mind the sense of vision.

64. The figure represents a horizontal section of the human eye; its general form is spherical, but the front BAB' is more convex than the rest.



The outer coat is called the *sclerotica*, and is white and opaque, except in the front, which is occupied by the transparent convex portion BAB' called the *cornea*.

The interior of the *sclerotica* is lined with a soft thin coat, called the *choroid membrane;* at the junction of this membrane with the sclerotica arises the *uvea*, an opaque membrane, having an aperture aa' in its centre called the *pupil*, through which light enters the eye, and by the spontaneous enlargement and diminution of which, the quantity of light admitted is regulated. The diameter of the pupil varies from about $\frac{1}{10}$th to $\frac{1}{4}$th of an inch.

The interior of the choroid membrane is covered with a black substance called the *pigmentum nigrum*, the office of which is to absorb stray rays of light, and so prevent internal reflexions, which would be the source of much confusion. At the back of the eye, and imbedded in the pigmentum

nigrum, is the *retina;* the retina is a network of very fine nerves, which branch off from, and may be looked upon as a continuation of, the optic nerve, which proceeding directly from the brain enters at O, on the side of the eye next the nose.

CC' is a soft, transparent substance, in the form of a double convex lens, and called the *crystalline lens.*

The space between the crystalline lens and the cornea, is filled with a transparent fluid resembling water, and called the *aqueous humour.* That between the crystalline and the retina is filled with another humour, called the *vitreous humour.* The refractive indices of these humours differ very little from that of water.

65. When a pencil of light diverges from a luminous point upon the exterior surface of the eye, it suffers refraction at the cornea, and again at the surface of each successive humour through which it passes, and by the combined refraction of all is made to converge to a point upon, or very near to the retina. In like manner the image of an external object is formed upon the retina; it is easy to see that the image will be inverted with respect to the object.

66. The focal length of the eye is not constant, but is varied instinctively by the eye, so as to adapt itself to vision at different distances. The shortest distance to which the eye can adjust itself varies in different persons, and is called the *least distance of distinct vision;* with regard to vision of distant objects, the majority of persons can see when the object is at such a distance that the rays from it may be considered to be parallel to each other, but some eyes require the rays entering them to have a certain degree of divergency: in general, however, we say that rays of light are fitted to produce distinct vision, when they are parallel to each other. It may be observed, that no eye can see by means of rays having the smallest amount of convergency.

67. The eyes of animals, though agreeing in principle, are different in many of their details from those of men; the

difference being, in general, such as can be accounted for, by consideration of the different circumstances of vision, to which it is desirable that they should be adapted: but the student, who desires information on this head, must consult other works on the subject.

ON DEFECTS OF SIGHT.

68. Perfect vision requires that a pencil of rays incident on the cornea should be made, by refraction through the several humours of the eye, to converge accurately to a point upon the retina. Vision will therefore be imperfect, when the rays converge to a point in front of the retina and then diverge upon it, or when they converge to a virtual focus behind it. The resulting defect is nearly the same in the two cases, for in both the image of a point, instead of being a point, is a small circle or disk of light, and the image of an object is therefore formed of such small circles, which, overlapping each other, produce indistinctness or confusion.

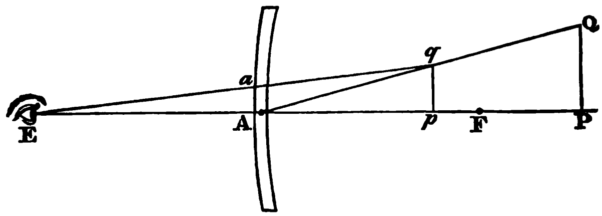
The former defect is that of *shortsight*, and arises from the too great convexity of the refracting surfaces of the eye. It may be remedied by the use of a *concave* lens, which will give the rays the degree of divergency necessary to enable the eye to bring them to a focus upon the retina.

The latter defect is that of *longsight*, and arises from the too great flatness of the refracting surfaces of the eye; it is a defect usually brought on by old age. It may be remedied by the use of a *convex* lens, which will cause the rays to enter the eye in a state of parallelism, and so to be fit to produce distinct vision.

ON VISION THROUGH A SINGLE LENS.

69. *To determine the angle under which a given object will be seen by the eye, when viewed through a concave lens.*

Let A be the centre of the lens, E the eye, PQ the object; join QA, and let pq be the image of PQ. Join Eq; then



qEp is the angle under which the object is seen through the lens, or the *visual angle*.

Let $PQ = \lambda$, $AP = u$, $AE = d$, $f =$ the focal length of the lens, $\theta =$ the visual angle.

Then $\tan\theta = \dfrac{pq}{AE + Ap} = \dfrac{PQ}{AE + Ap}\,\dfrac{Ap}{AP}$, by similar triangles,

$$= \frac{\lambda}{u}\,\frac{1}{1 + \dfrac{d}{Ap}} = \frac{\lambda}{u}\,\frac{1}{1 + d\left(\dfrac{1}{u} + \dfrac{1}{f}\right)}. \text{ (Art. 41.)}$$

Let α be the angle under which the object would have been seen by the naked eye, then $\tan\alpha = \dfrac{\lambda}{u + d}$;

$$\therefore \quad \frac{\tan\theta}{\tan\alpha} = \frac{u + d}{u + d + \dfrac{du}{f}}.$$

This ratio is less than unity, and the effect is that the object appears *diminished*.

If we had taken the case of a convex lens, we should have found, that

$$\frac{\tan\theta}{\tan\alpha} = \frac{u + d}{u + d - \dfrac{du}{f}}.$$

If $u + d$ be greater than $\frac{du}{f}$, the object will appear *magnified*, and it may be magnified to any extent. But if $\frac{du}{f}$ be greater than $u + d$, $\tan\theta$ will be negative, and the image will be inverted and not necessarily magnified.

70. If we suppose the lens so adjusted that the rays enter the eye in a state of parallelism, we must have $u = f$, and then (for the convex lens),

$$\frac{\tan\theta}{\tan\alpha} = \frac{f + d}{f + d - d} = 1 + \frac{d}{f}.$$

The magnification by a convex lens will be therefore greater as $\frac{1}{f}$ is greater; hence it has been proposed, to call the quantity $\frac{1}{f}$ the *power* of a lens.

71. A very small convex lens, of short focal length, or a very small sphere of glass, may be used as a magnifying glass in a way slightly differing from the preceding. When an object is placed very near the eye, a magnified image is formed on the retina, but on account of the too great divergency of the rays the eye is not able to obtain a distinct perception of the object. If now a very small lens, not exceeding in breadth that of the pupil of the eye, and of focal length so short that the object shall be in its principal focus, be placed close to the eye, the rays of light emerging from the lens will be parallel to each other, and therefore fit to produce distinct vision, and the magnification of the image on the retina will be the same as before. The rays will in this case pass centrically through the lens.

ON THE GENERAL PRINCIPLE OF TELESCOPES.

72. When a small object at a great distance is viewed by the naked eye, there are two reasons why the vision is indistinct, namely, the smallness of the angle which the object subtends at the eye or the *visual angle*, and the small quantity of light which comes to the eye from any point of the object. The ends to be accomplished therefore by an instrument used for viewing distant objects are also two, namely, to increase the visual angle, and to increase the quantity of light which reaches the eye. The latter is accomplished by allowing the rays to fall upon a convex lens, called the object-glass, which collects from each point of the object a quantity of light, bearing to the quantity which would enter the naked eye, the ratio of the area of the object-glass to the area of the pupil of the eye; and the former by deflecting the rays through a system of lenses, the arrangement of which varies in telescopes of different constructions. In some telescopes the rays are received on a concave reflector, instead of a convex lens, but the principle is the same.

73. It may be seen, without any difficulty, that the two defects of vision, which are to be remedied, are in some measure antagonistic. For, suppose, that from a very small distant object a certain quantity of light falls on the object-glass of a telescope, then the magnifying power is greater in proportion as the light is spread over a *larger extent* of the retina; but the brightness of the illumination of the retina is greater in proportion as the light is more concentrated, or spread over a *smaller extent* of the retina; hence, when an image is formed on the retina, we have this general relation,

$$\text{brightness of the image} \propto \frac{1}{\text{magnification produced}}.$$

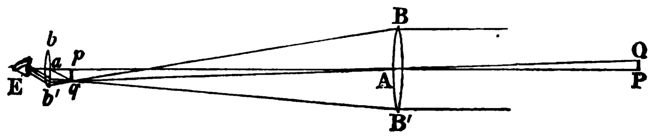
If therefore, with a given object-glass, we increase the magnifying power, we diminish the brightness of the image, and thus we have a limit put to increase of magnification. Hence, in order that we may have telescopes of great power, we must have object-glasses, or object-mirrors, of very large diameters, and in the construction of these lies the difficulty of making powerful telescopes.

74. We shall now proceed to describe the construction of various telescopes, premising that we shall describe them in their simplest form; to render them practically useful instruments, it would be necessary to introduce a variety of refinements of construction, which the elementary mode of treating the science of Optics adopted in this treatise precludes us from making intelligible to the student.

Telescopes may be divided into two classes, Refracting, and Reflecting.

ON REFRACTING TELESCOPES.

75. *The Common Astronomical Telescope.*



The common astronomical telescope consists of two convex lenses, a larger one BAB' called the object-glass, and a smaller one bab' called the eye-glass, placed on the same axis, and at a distance from each other equal to the sum of their focal lengths.

Let the axis of the instrument be directed to a point P of a very small object PQ, so distant that rays from any

point of it which fall upon the object-glass are sensibly parallel; then an inverted image pq will be found in the focus of the object-glass, and the rays which diverge from any point q of the image upon the eye-glass will, after refraction, emerge approximately parallel to the line qa, which joins q with the centre of the eye-glass, since aq nearly $= ap$ = the focal length of the eye-glass. If therefore an eye be placed at E, the point at which the axis of the pencil from q crosses the axis, the rays entering the eye will be fit for producing distinct vision, and an image of PQ will be seen.

Objects seen through this telescope will appear inverted, but this is of no importance in the case of celestial objects.

76. *The Magnifying Power.*

The magnifying power is measured by the ratio of the visual angles, when the object is viewed through the telescope and with the naked eye respectively.

The angle under which PQ would be seen with the naked eye is $PAQ = pAq$; and the angle under which pq is seen is paq, since the rays emerge parallel to qa;

$$\therefore \text{ magnifying power} = \frac{paq}{pAq} = \frac{Ap}{ap} \text{ nearly,}$$

$$= \frac{f_o}{f_e};$$

where f_o, f_e represent the focal lengths of the object-glass and eye-glass respectively.

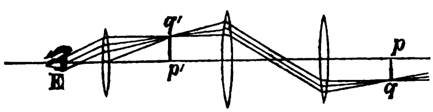
77. *The Field of View.*

The field of view is the angle subtended at the eye, or at the centre of the object-glass, (for on account of the distance of the object the two will be sensibly the same,) by the largest object which at a given distance can be seen through the telescope. This definition, however, though apparently precise, is not so really; for, suppose we find a point in the object

from which a pencil of rays, after being refracted through the object-glass, just fall on the eye-glass: then if we take a point a little further from the axis, the rays from it will not all fall on the eye-glass, but some of them will be lost; still more will this be the case with a point a little further from the axis; and so on, until at last we come to a point from which no rays fall upon the eye-glass, and therefore none reach the eye. The result is, that in looking through a telescope such as we have described, the outer portions of the field instead of being clearly defined would gradually fade away; this imperfect part of the field is called *the ragged edge of the field of view.*

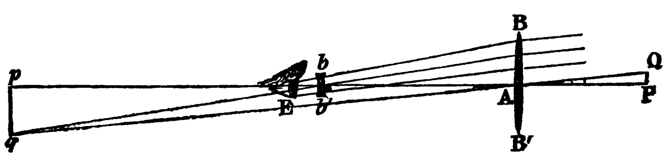
The ragged edge may be cured by placing a *stop*, or annulus of metal, in the focus of the object-glass; for by this means we can stop any rays we please, and limit the field to any extent. If this be done, the angular extent of the field of view will be the angle subtended by the aperture of the stop at the centre of the object-glass.

78. The telescope which we have now described is not applicable to vision of terrestrial objects, on account of its inverting property; but it may be adapted to the purpose by using a combination of lenses called *an erecting eye-piece*, instead of the simple eye-glass. The construction of such an



eye-piece will be sufficiently understood from inspection of the annexed figure, which represents one form of its construction.

79. *Galileo's Telescope.*



This telescope consists of a convex lens BAB', and a concave lens bab', placed on the same axis, at a distance from each other equal to the difference of their focal lengths; BAB' is the object-glass, and is of much greater breadth and focal length than the eye-glass bab', which need not be larger than the pupil of the eye.

Let the axis of the instrument be directed to a point P in an object PQ, which is at such a distance that rays from any point of it may be considered to fall upon the object-glass in a state of parallelism; then if nothing were interposed, an inverted image pq would be formed of PQ in the focus of the object-glass. If now an eye were placed at E the rays converging to any point q would not produce distinct vision (see Art. 66); but if a small concave lens bab' be placed before the eye, and at a distance from the image equal to its focal length, the rays will emerge in a state of parallelism, and therefore will produce distinct vision; and the visual angle will be paq, since the rays which before refraction at the eye-glass were converging to q emerge parallel to aq. The rays being intercepted by the eye-glass before they have crossed the axis, objects will appear erect.

80. *The Magnifying Power.*

Let f_o, f_e be the focal lengths of the object-glass and eye-glass respectively; then the visual angle for PQ seen with the naked eye is PAQ, the visual angle when seen through the telescope is paq;

$$\therefore \text{ magnifying power } = \frac{paq}{PAQ} = \frac{paq}{pAq} = \frac{f_o}{f_e} \text{ nearly.}$$

81. *The Field of View.*

The ragged edge in Galileo's telescope is not curable by the use of a stop as in the astronomical telescope, because no real image is allowed to be formed by the object-glass, and it is manifest therefore that a stop placed any where within the telescope will not stop imperfect pencils only.

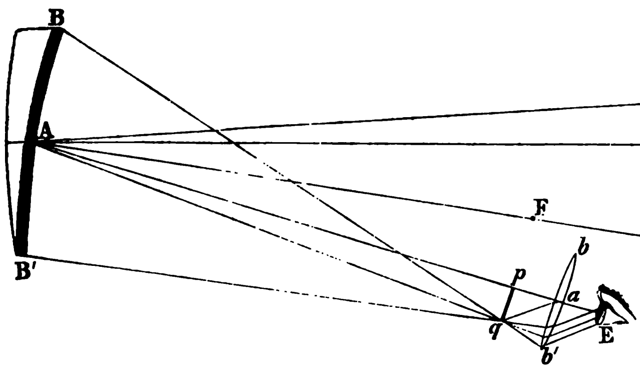
In this telescope the field of view will be limited by the object-glass, and not by the eye-glass, as in the case of the astronomical telescope. For the field will necessarily be limited in any combination of lenses, by the first lens through which the rays pass excentrically; and in Galileo's telescope, the rays pass excentrically through the object-glass; for although rays fall upon the whole extent of the object-glass from each point of the object, yet the eye-glass selects a small pencil, (as will be seen distinctly from the figure,) and this small pencil, with which alone we are concerned, passes through the object-glass excentrically.

82. Galileo's telescope possesses great historical interest, as being the first combination of lenses so used; the construction is, however, now only applied to opera-glasses, and for astronomical purposes is wholly useless. The capital defect of the telescope is that no image is formed by the object-glass; now observations of the stars are made by means of fine wires, which, being placed in the focus of the object-glass of an astronomical telescope, become visible by stopping pencils of rays which there converge to points, and the place of a star is noted by referring it to these wires; but in Galileo's telescope wires cannot be so used, there being no position in which they can be made visible; hence for astronomical purposes the construction is totally useless.

ON REFLECTING TELESCOPES.

83. In reflecting telescopes, the place of the object-glass is supplied by an object-mirror, or concave speculum of metal, which reflects the incident rays and causes them to converge. We shall describe four kinds of reflecting telescopes, which involve however, (as will be seen) only two essentially different constructions.

84. *Herschel's Telescope.*



BAB' is a concave speculum, having its axis *AF* inclined at a small angle to *AP* the axis of the tube of the telescope, so that a ray incident in the direction of the axis of the tube is reflected in the direction *Ap*; on the line *Ap* as axis is placed the convex lens *bab'*, at a distance from *A* equal to the sum of the focal lengths of the mirror and lens.

If the axis of the instrument be directed to *P*, a point in a small distant object *PQ*, an inverted image *pq* will be formed on *Ap*, and at a distance from *A* equal to the focal length of the mirror: and this image being by the construction in the focus of the eye-glass, the rays after refraction through it will emerge in a state of parallelism, and will therefore be fit for the production of distinct vision. Objects will appear inverted.

The form of the object-mirror should be parabolical, not spherical. See Art. 25.

85. This is the most simple construction of the reflecting telescope, and is quite analogous in every respect to that of the astronomical telescope. It is, however, adapted only for very large instruments, because if the telescope be not large, the observer's head, when looking into the telescope, will materially interfere with the incidence of the rays: this

defect is obviated by Newton's construction, which we shall presently describe.

86. *The magnifying power.*

The visual angle for the object PQ viewed with the naked eye is PAQ; when viewed through the telescope it is paq.

$$\therefore \text{ magnifying power } = \frac{paq}{PAQ} = \frac{paq}{pAq} = \frac{Ap}{ap}, \text{ nearly,}$$

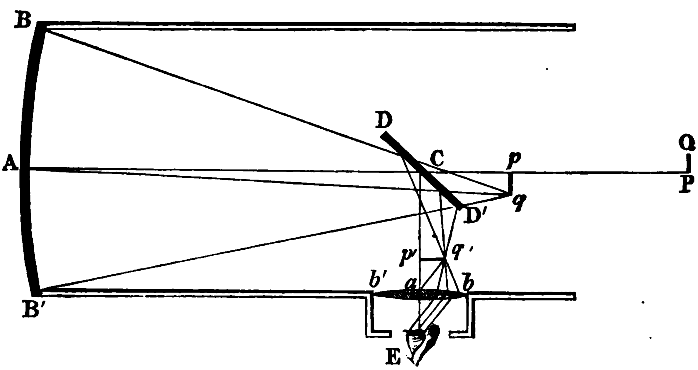
$$= \frac{f_o}{f_e},$$

where f_o f_e are the focal lengths of the object-mirror and eye-glass respectively.

87. *The field of view.*

The ragged edge may be cured, as in the astronomical telescope, by placing a stop at the common focus of the mirror and eye-glass, and the field of view will then be measured by the angle which the diameter of the stop subtends at the central point A of the mirror. If there be no stop, we may take the angle subtended by the eye-glass at the same point as the measure of the field of view.

88. *Newton's Telescope.*

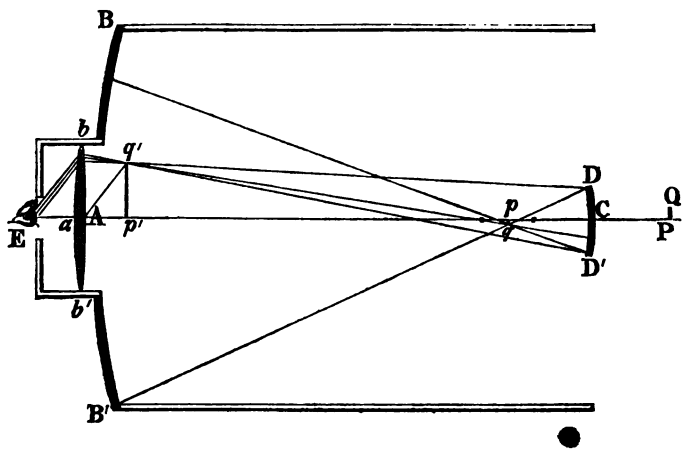


BAB' is a concave mirror, which, if light were incident upon it from a small distant object *PQ*, would form an inverted image *pq* of *PQ* in the principal focus. But a small plane mirror *DCD'*, placed at an angle of 45^0 with the axis of the object-mirror, causes the image to be formed at *p'q'*, instead of *pq*, and if a convex lens *bab'* be placed on *Cp'* as axis, and at a distance from *p'q'* equal to its focal length, the rays will emerge after refraction through *bab'* in a state of parallelism, and therefore a distinct image of *PQ* will be seen by an eye at *E*.

89. *The magnifying power, and field of view.*

Both of these will be the same as in Herschel's construction; the two telescopes are in fact the same, the plane mirror in Newton's being introduced principally for the sake of avoiding the necessity of looking directly towards the object-mirror, which in the case of small instruments would manifestly be most inconvenient.

90. *Gregory's Telescope.*

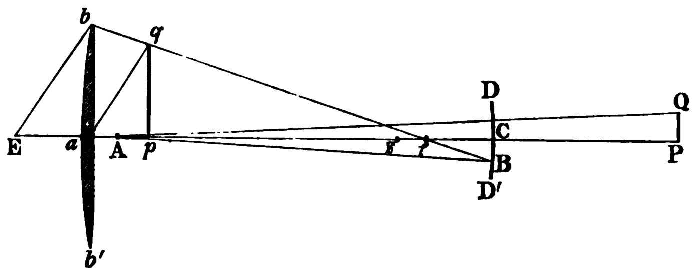


BAB' is a concave mirror, which being directed to a distant object *PQ* forms an inverted image of it *pq* in its principal focus. *DCD'* is a small concave mirror, on the same axis as

BAB′, and at a distance from *p* rather greater than its focal length; the rays from the different points of *pq* diverge upon *DCD′*, and after reflexion form an inverted image *p′q′* of *pq*, which will therefore be erect with respect to *PQ*. The adjustment is such that *p′q′* is formed in the focus of a convex lens *bab′*, on the same axis as the two mirrors, and on which the light falls through a circular aperture in the middle of the object-mirror; the rays of light therefore, after refraction through it, emerge in a state of parallelism, and produce distinct vision of the object *PQ* to an eye at *E*.

Objects seen through this telescope will appear erect.

91. *The magnifying power.*



To find the magnifying power let us trace the course of the ray, which *would* fall, from the point *Q* of a distant object, on *A* the middle point of the object-mirror; let *AB* be the reflected ray, which, after reflexion at the small mirror will pass very nearly through *f* the focus of the small mirror, (on account of the distance of the point *A* being very considerable,) and passing through the point *q* of the final image, and being refracted through the eye-glass, emerges in the direction *bE*, parallel to *qa*.

Then *QAP* may be regarded as the visual angle of *PQ* to the naked eye, *qap* the visual angle when the object is viewed through the telescope;

therefore magnifying power $= \dfrac{qap}{QAP} = \dfrac{qap}{qfp}\dfrac{BfC}{BAC} = \dfrac{fp}{ap}\dfrac{AC}{fC}$

$$= \frac{f_o}{f_e}\frac{f_o + f_m}{f_m} \text{ nearly,}$$

where $f_o f_m f_e$ are the focal lengths of the object-mirror, small mirror, and eye-glass respectively.

92. *Field of view.*

The field of view may be limited either by the eye-glass, or by the small mirror. With the same figure as in the last article, suppose the ray there traced to fall on the extreme point b of the eye-glass, then the eye-glass will limit the field, and we shall have,

the field of view $= 2PAQ = 2BAC$

$$= 2\frac{fC}{AC}.BfC = 2\frac{fC}{AC}.afb$$

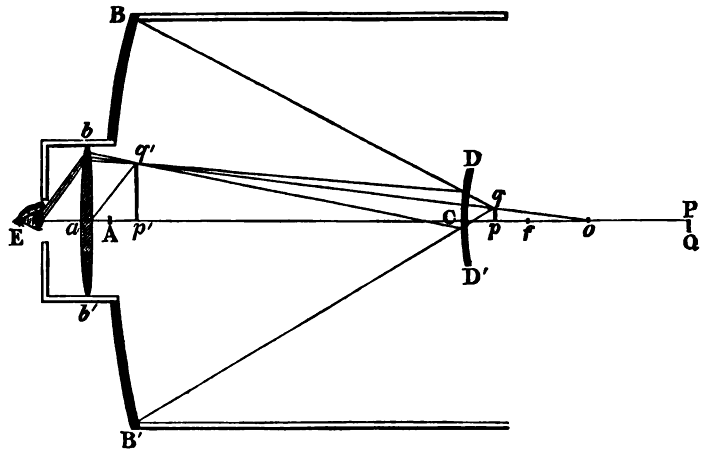
$= \dfrac{f_m}{f_o + f_m}\dfrac{b}{f_o}$ nearly, where b is the breadth of the eye-glass.

If the field of view is limited by the small mirror, we must suppose the ray traced in the figure to pass through the extreme point D of the small mirror, in which case the field of view $= \dfrac{DD'}{AC} = \dfrac{c}{f_o + f_m}$, where c is the breadth of the small mirror.

In order that the field, as limited by these two considerations, may be the same, we must have

$$\frac{c}{b} = \frac{f_m}{f_o}.$$

93. *Cassegrain's Telescope.*



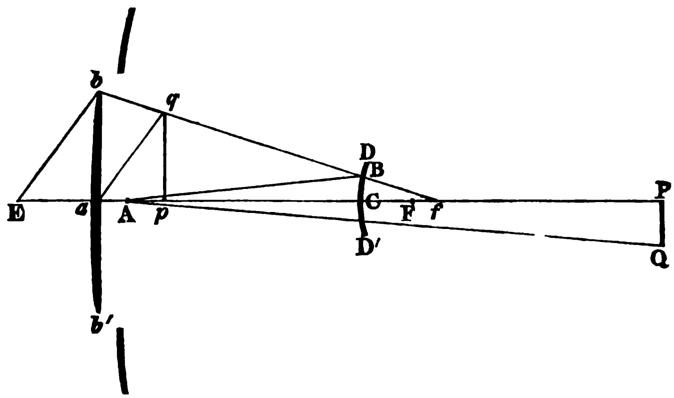
The construction of this instrument differs from that of the preceding, only in having a convex small mirror instead of a concave.

BAB' is a concave mirror, which, being directed to a distant object PQ, would form an inverted image of it pq in its principal focus; but the reflected rays are intercepted by the small convex mirror DCD', which is so placed that its focus is a little further from the object-mirror than the principal focus of that mirror, and consequently an image $p'q'$, inverted with respect to PQ, is formed at some distance from the small mirror. The adjustment is such that this image is formed in the focus of the eye-glass bab', and consequently the rays after refraction emerge in a state of parallelism, and are therefore fitted to produce distinct vision.

Objects seen through this telescope will appear inverted.

94. *The magnifying power.*

Let a construction be made similar to that for Gregory's telescope.



Then the magnifying power $= \frac{qap}{QAP} = \frac{qap}{qfp} \frac{BfC}{BAC}$

$$= \frac{fp}{ap} \frac{AC}{fC} = \frac{f_o}{f_e} \frac{f_o - f_m}{f_m} \text{ nearly.}$$

95. *Field of view.*

If the field be limited by the eye-glass we shall have,

$$\text{field of view} = 2PAQ = 2BAC$$

$$= 2\frac{fC}{AC} BfC$$

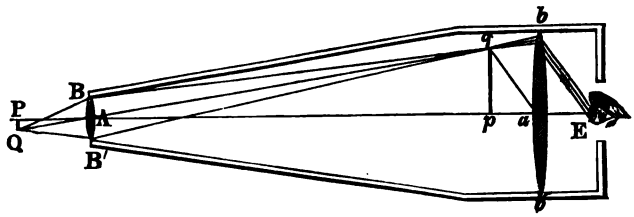
$$= \frac{f_m}{f_o - f_m} \frac{b}{f} \text{ nearly.}$$

It will be observed that the expressions for the magnifying power and field of view may be obtained from those in Gregory's construction, by changing the sign of f_m.

96. *The Compound Refracting Microscope.*

We have already explained (Art. 71) the principle upon which a small convex lens of very short focal length may be used as a magnifying glass, or simple microscope. Combinations of lenses may be used for the same purpose, or combi-

nations of reflectors and lenses, and such combinations are called *compound* microscopes. We shall confine ourselves to the description of the compound refracting microscope in its simplest form, observing that to make it practically useful a number of refinements must be introduced.



BAB′ is a small convex lens, before which, and at a distance from it a little greater than its focal length, if a small object *PQ* be placed, an inverted image *pq* will be formed of it. The adjustment is such that *pq* is formed in the focus of a convex lens *bab′*, and therefore the rays when refracted through it emerge in a state of parallelism, and therefore in a state fit to produce distinct vision; and an eye at *E* will see a magnified inverted image of *PQ*.

ASTRONOMY.

ASTRONOMY.

1. We propose in the following treatise to give some account of the physical constitution of the universe, the motions of the heavenly bodies, and the resulting phenomena, with the mode of making observations, all which and other kindred subjects are classed under the head of Plane Astronomy; we shall not here treat of the Physical branch of astronomy, which investigates phenomena on the principles of Mechanics, and refers them to general laws, but only of that branch which deals with facts as matters of observation.

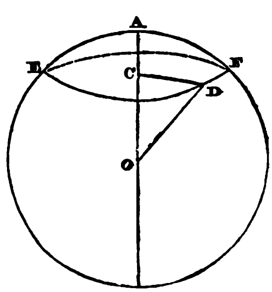
As it is of the utmost importance that the student should be perfectly familiar with the notion of a *sphere*, and of lines drawn upon it, we shall commence by presenting him with a few of the most elementary propositions and notions belonging to the doctrine of the sphere.

ON THE SPHERE.

2. Def. A sphere is a surface every point in which is equidistant from a given point, called its *centre*.

The distance from the centre to the surface is called the *radius*, and any line passing through the centre and bounded by the surface is called a *diameter* of the sphere.

3. *Every section of a sphere made by a plane is a circle.*



Let EDF be any such section of a sphere, of which the centre is O. Draw OC from O perpendicular to the cutting plane, and join CD, OD, D being any point in the section EDF.

Then since OC is perpendicular to the cutting plane, it is perpendicular to any line in that plane, and therefore to CD;

$$\therefore \quad OD^2 = OC^2 + CD^2,$$

$$\text{or} \quad CD^2 = OD^2 - OC^2;$$

but OD and OC are both constant quantities, therefore CD is constant, or the section is a circle having C for its centre.

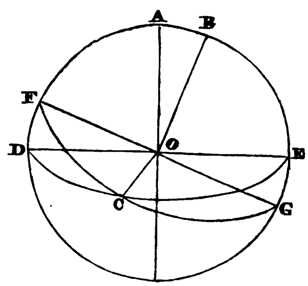
4. A section of a sphere made by a plane passing through the centre is called a *great* circle; other sections are called *small* circles.

The diameter of the sphere, which is perpendicular to the plane of any circle on the sphere, is called the *axis* of that circle; and the points in which the axis meets the sphere are called the *poles* of the circle.

It is evident that the poles of a great circle are equidistant from it; also it is easy to see that the pole of a circle is equidistant from every point in the circle.

5. *The angle which is subtended at the centre of a sphere by the arc joining the poles of two great circles, is the angle of inclination of the planes of the circles.*

Let O be the centre of the sphere; DCE, FCG the two great circles; OA, OB lines respectively perpendicular to the planes of the circles, so that A, B are their poles. Join OC.



Then OC being in the plane DCE is perpendicular to AO, and being in the plane FCG, is perpendicular to BO; therefore CO is perpendicular to the plane in which AO, BO lie, and therefore to EO and GO;

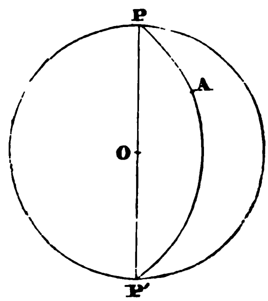
$\therefore$ EOG is the inclination of the planes of the circles.

But $EOG = 90^0 - BOE = AOB$;

therefore the angle subtended by AB is the inclination of the planes of the circles.

COR. Hence also it appears, that the arc joining the points in two great circles distant 90^0 from the point of their intersection, subtends at the centre of the sphere an angle equal to the inclination of the planes of the circles.

6. *To determine the position of a point on a sphere.*



Let A be a point on a sphere, the centre of which is O; then its position may be most conveniently determined as follows.

Let POP' be a given diameter of the sphere; through P A and P' draw the great circle PAP'; then if the angle which the plane of PAP' makes with a given plane passing through PP', and the arc PA, be given, the position of A will be completely determined.

ON THE FIGURE OF THE EARTH.

7. The form of the earth is nearly, but not accurately, spherical. Its true form is that of a slightly oblate spheroid, or a surface generated by the revolution of an ellipse, having its axes nearly equal, about its minor axis. In the greater number of cases it is sufficient to consider the earth's figure to be that of a sphere.

The round form of the earth is easily concluded from such considerations as the following; the top of the masts are the first portion of a ship which become visible; all the heavenly bodies with which we are acquainted have that form; and moreover the earth has been actually sailed round. The experiments which determine the actual figure to be spheroidal are of a far more delicate kind, and cannot be entered upon here.

8. The heavens present to an observer on the earth's surface the appearance of a hollow sphere, at the centre of which the observer stands, and it will be convenient to conceive of such a sphere which we may call the *celestial sphere*, and in the surface of which we may conceive the heavenly bodies to be; the actual point on the surface of the celestial sphere to which we shall refer any given object, will be the point in which the line joining the eye of the observer, and the object meets the surface of the sphere.

On account of the enormous distance of the heavenly bodies from the earth, it will be, for many purposes, indifferent whether we consider the centre of the earth or the position of the observer as the centre of the celestial sphere.

9. The earth revolves, as will be explained more particularly presently, about a certain axis coinciding very nearly with the minor axis of its figure, considered as a spheroid; the points in which this axis produced meets the celestial sphere are called the *North and South Poles*. The great circle of which these points are the poles is called the *equator*, and the two equal portions into which the plane of the equator divides the celestial sphere are called the *Northern and Southern Hemispheres*.

The plane of the equator cuts the surface of the earth into two equal portions, which are also called respectively the northern and southern hemispheres; and the circle in which the plane cuts the earth is sometimes called the equator, as well as that in which it cuts the celestial sphere.

The direction of a line perpendicular to the surface of still water at any place on the earth's surface, is called the *vertical* at that place; and the points in which the vertical line meets the celestial sphere, are called the *zenith* and *nadir* of the place. The vertical direction will be very approximately that of the line joining the place with the earth's centre.

A plane perpendicular to the vertical at the earth's surface, is called the *sensible horizon;* a plane perpendicular

to the same line at the earth's centre, the *rational horizon.* In the greater number of cases, the sensible and rational horizon may be considered as coincident.

The *meridian* of a place is the great circle passing through the poles of the heavens and the zenith of the place.

ON TERRESTRIAL LATITUDE AND LONGITUDE.

10. The position of a place upon the earth's surface may be determined on the principle explained in Art. 6. Let the meridian of some place, as Greenwich, be considered as given; then the angle between the meridian of Greenwich and that of the place in question, is called the *longitude* of the place, and the angle subtended by the arc of the meridian between the zenith of the place and the equator, is called the *latitude;* and the latitude is said to be *north* or *south,* according as the place is to the north or south of the equator. The latitude and longitude being given, the position of the place is defined.

The complement of the angle which measures the latitude of a place, is called the *co-latitude.*

It is usual to measure longitude through 180^0 east and west of Greenwich; perhaps it would be more convenient to measure through 360^0 in the same direction.

ON THE EARTH'S MOTION.

11. The motion of the earth may be conceived of, as being compounded of two motions, namely, a motion of revolution about an axis, while at the same time that axis is moving in space.

Let us first consider the revolution about the axis: and for a first approximation we may say, that the earth revolves

about a line coinciding with its shorter axis and remaining fixed in space; this we shall find afterwards to be not strictly true. The time of revolution is twenty-four hours; and the effect produced to an observer on the earth's surface is this, that, imagining himself to be fixed in position, the celestial sphere appears to revolve about its poles, carrying the heavenly bodies with it; so that the sun and stars describe circles about the axis of revolution, the greater part small circles, those only describing great circles which happen to be in the equator. When a heavenly body comes into the horizon of any given place it is said to *rise*, when it reaches the meridian it *culminates*, and when it again reaches the horizon it *sets*.

All this coincides with observation; for the stars are observed to revolve about a certain point in the heavens, nearly coinciding with a bright star, known as the Pole Star, and we are therefore obliged to adopt one of two hypotheses, namely, that the celestial sphere actually revolves about the earth as fixed, or that the celestial sphere being fixed the earth revolves about an axis which remains fixed in space. The great simplicity of the latter hypothesis leaves no doubt concerning its truth.

This is the *diurnal* motion of the earth, which gives rise to the succession of day and night; in addition to this there is an *annual* motion, that is, the earth is carried round the sun in a certain period, which constitutes one year. In this motion the axis of revolution moves always parallel to itself, as is shewn by the fact of its appearing always to point to the same point of the celestial sphere. The centre of the earth does in fact describe an ellipse in one plane about the centre of the sun, and this ellipse does not differ much from a circle; at present, however, we are not concerned with the actual path described by the earth, but only with the fact of its moving round the sun in the plane of a great circle, in the course of a year. According to observation, the sun appears to move in that time round the earth, but the phenomena will be exactly the same, whether the earth move round the sun, or the sun round the earth, and the

consideration of the enormous magnitude of the sun as compared with the earth, combined with other reasons which will appear hereafter when we come to treat of the planets, leave no doubt as to the correctness of the hypothesis of the motion of the earth about the sun, not the sun about the earth.

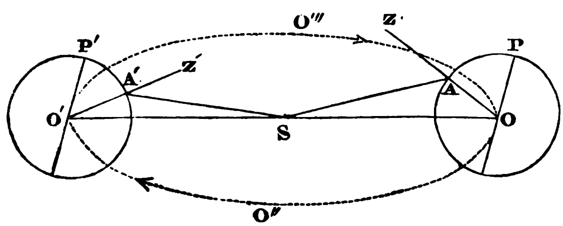
12. For purposes of explanation however, we shall in general speak of the sun as moving in a great circle about the earth, and this great circle we shall call the *ecliptic.*

The inclination of the plane of the equator to that of the ecliptic is an angle of about $23^{0}\,28'$, and is called the *obliquity of the ecliptic.*

ON THE SEASONS.

13. If a small plane be exposed to heat, which emanates from any given source, as for instance the sun, it is manifest that the quantity of heat received will be greatest when the plane is perpendicular to the direction in which the heat emanates, and the smaller the angle which the plane makes with that direction, in other words, the greater the *angle of incidence* of the sun's rays, the smaller will be the quantity of heat received. On this principle we can explain the change of the seasons.

14. Let $OO''O'O'''$ be the path of the earth's centre

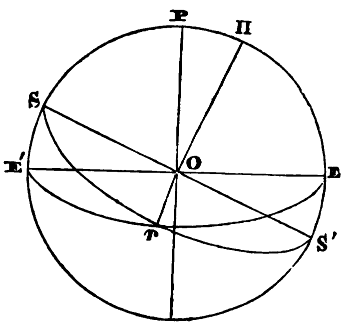


round the sun S. Then the earth revolves about an axis,

(OP or $O'P'$,) making an angle of $23^{\circ}28'$ with the perpendicular to the plane of this path.

Consider the position O of the centre of the earth when the angle SOP is the greatest possible, that is, when $SOP = 113^{\circ}28'$. Let A be any point on the earth's surface; join OA and produce it to Z, and join SA: then the angle of incidence of the sun's rays is SAZ. Now consider the exactly opposite position of the earth in its orbit, when the angle $SO'P'$ is the least possible, that is, when $SO'P' = 66^{\circ}32'$; and let A' be the position of the place A in this case; then joining $O'A'$ and producing it to Z', the angle of incidence of the sun's rays is $SA'Z'$, an angle much less than SAZ. Hence the heat in the position A' will be much greater than in the position A: A' will, in fact, be enjoying *mid-summer*, A will be in *mid-winter*. The positions O'' O''' will be positions of intermediate heat, and will correspond to *spring* and *autumn* respectively.

ON THE SUN'S MOTION IN THE ECLIPTIC.



15. Let O be the centre of the earth, supposed fixed; and let the plane of the paper pass through O, the pole of the equator P, and the pole of the ecliptic Π; $E \,♈\, E'$ is the equator, $S \,♈\, S'$ the ecliptic.

Then we may say, that the sun describes its course in the ecliptic round O uniformly in the course of a year. The ecliptic is conceived to be divided into twelve equal portions, each therefore consisting of 30°, and these portions are called the *twelve signs of the Zodiac;* they are known by the following names: Aries, Taurus, Gemini, Cancer, Leo, Virgo, Libra, Scorpio, Sagittarius,

Capricornus, Aquarius, Pisces; the origin of these names will be seen hereafter; they are denoted by different symbols, one only of which we shall use, (♈), which is the symbol for *Aries*. The first point of Aries is determined by the intersection of the equator and ecliptic, the other point of intersection being the first point of Libra.

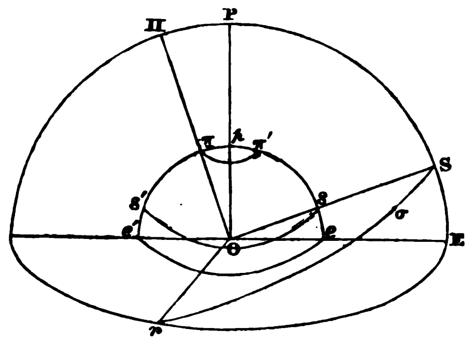
Suppose the sun to be at ♈, which will happen at the time of year called the *vernal equinox*, and suppose it to be moving in the ecliptic towards S: for three months its distance from the equator will increase, and at the end of that time it will be at S, a point in the great circle passing through P and Π. Its distance from the equator will then diminish, until at the end of three months it will again be in the equator; this will happen at the *autumnal equinox*. The sun will now go to the south of the equator, and at the end of the next three months will be at S' on the great circle passing through P and Π. Lastly, after passing S' the sun will again approach the equator, and at the end of three months more will be again at ♈.

16. For a few days before and after passing the points S and S', the sun will move nearly parallel to the equator, and therefore, will neither approach it nor recede from it; hence, so far as motion to or from the equator is concerned, the sun may be said at those points to be stationary for a short period, and they are on this account called the *solstices*. S is the *summer* solstice, S' the *winter* solstice; the great circle passing through the solstices and the poles of the equator and ecliptic, is called the *solstitial colure*.

ON CLIMATE.

17. It will be easily seen, that all parts of the earth's surface are not equally affected by the sun's heat; the term *climate* is used to express the difference between the several regions of the earth in this respect.

Let O be the centre of the earth; ♈ S the ecliptic, ♈ E the equator, Π, P, their respective poles; and let π, p, s, e,



be the points in which the lines $O\Pi$, OP, OS, OE, respectively meet the earth's surface; also $\pi\pi'$, ss', small circles on the earth's surface made by planes parallel to the equator, and ee' the great circle in which the plane of the equator cuts the earth.

Then the portion of the earth's surface to the north of $\pi\pi'$, and an equal portion to the south of a similar small circle round the south pole, are called the *frigid Zones:* the portion between $\pi\pi'$ and ss', and a similar portion in the southern hemisphere, are called the *temperate Zones:* and the portion between ss' and a similar circle in the southern hemisphere is called the *torrid Zone.*

The small circle $\pi\pi'$ is called the *Arctic Circle,* and a similar one in the southern hemisphere the *Antarctic Circle.* The small circle ss' is called the *Tropic of Cancer,* and a similar one in the southern hemisphere the *Tropic of Capricorn;* because the sun after receding from the equator, turns (τρεπει) on entering the signs of Cancer and Capricorn, and again approaches the equator.

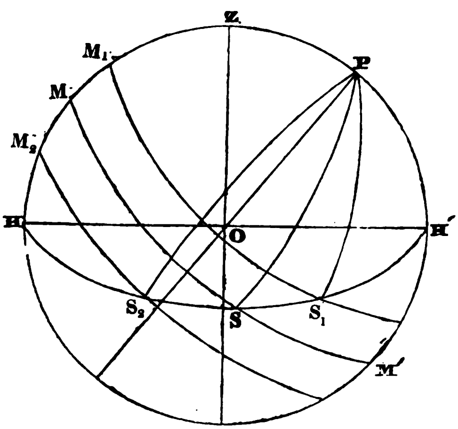
18. The peculiarity of the torrid zone is that a place situated within it will have the sun *vertical,* that is, exactly in its zenith, twice in the year. For let σ be the place of the

sun at any given time, then, if we join $O\sigma$, this line will manifestly intersect the earth's surface at some point between the tropics, and that place, with all others in the same latitude, will have the sun in the zenith; and the same thing will take place when the sun has passed the solstice by a distance equal to σS. Even when the sun is not exactly vertical, its rays fall with a smaller obliquity on places between the tropics than in the temperate zones; hence the extreme heat of tropical climates.

The peculiarity of the frigid zones will be noticed in Article 20.

ON THE LENGTH OF THE DAY.

19. The motion of the sun in the ecliptic during one day is not very great; hence in considering the effect of its motion on the length of the day it will be sufficient to suppose it to preserve the same position in the ecliptic during one revolution of the earth, or during one of its own apparent revolutions about the earth.



Let O be the earth's centre, Z the zenith of a place on its surface, HH' the horizon of the place, MSM' the equator,

P its pole. Then we can determine the length of the day for any given position of the sun in the ecliptic, by supposing it to describe a small circle in a plane perpendicular to OP; as long as it is above the horizon it is day, the remainder of the twenty-four hours is night. We shall consider three cases.

(1) Let the sun be in the equator; then its diurnal path will be the great circle SM, and if we join PS by an arc of a great circle, half the day will be measured by the angle MPS (called an *hour-angle*). But it is not difficult to see that the angle MPS is a right angle, hence the hour-angle measuring the length of the day is two right angles; consequently that which measures the length of the night must be two right angles, or the day and night are equal. When therefore the sun is in the equator, day and night are equal all over the world; hence the equator is sometimes called the *equinoctial line*, and the first points of Aries and Libra are called the *equinoxes*.

(2) Suppose the sun to be in the summer solstice; then its diurnal path will be the small circle S_1M_1, the arc MM_1 being that which measures the obliquity of the ecliptic. Join S_1P by an arc of a great circle; then the hour-angle S_1PM_1 measures half the day, and this angle is greater than a right angle, hence the days are longer than the nights to places in northern latitudes.

(3) In like manner, if the sun is in the winter solstice, and S_2M_2 its diurnal path, the hour-angle S_2PM_2 will measure half the day, and the days will be shorter than the nights.

For intermediate positions of the sun the results will be easily inferred.

20. Let us consider the peculiarities of day and night in the frigid zones. Let us suppose PZ to be equal to the obliquity of the ecliptic, then the small circle S_1M_1 will pass through H', that is to say, when the sun is in the summer solstice it just does not set to a place upon the Arctic Circle, and in general, in order that the sun may set to any given

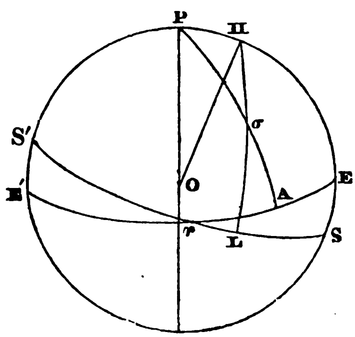
place its angular distance from the Pole, (or *North Polar distance*, as it is called,) must be greater than the latitude of the place; suppose, for instance, we take a place in latitude 70^0, then the sun will not set to that place from the time that its north polar distance is 70^0, until after having passed the solstice its north polar distance is 70^0 again. Corresponding to those long summer days there will be equally long winter nights; and at the poles there will be a day of six months in length, succeeded by a night equally long.

21. The greater length of day in summer than in winter, is an additional reason for the greater heat of that season; for since the earth receives heat from the sun during the day, and gives back the heat by radiation from its surface, the longer time the sun is above the horizon the greater will be the quantity of heat received by a body on the earth's surface, both directly from the sun, and also by radiation from the earth.

ON THE MODE OF DETERMINING THE PLACE OF A HEAVENLY BODY.

22. This will be done on the general principle explained in Art. 6, and already applied to the case of terrestrial latitude and longitude in Art. 10.

Let O be the centre of the celestial sphere, $E \Upsilon E'$ the equator, $S \Upsilon S'$ the ecliptic, $P \Pi$ their respective poles. Let σ be any heavenly body, the position of which we desire to determine: draw through σ the arcs of great circles $P\sigma A$, $\Pi\sigma L$; then ΥA is called the *Right Ascension*, $A\sigma$ the *Declination* of σ, and if these be given the position of σ will be determined. It will be the same thing if we suppose $P\sigma$,



the North Polar distance, to be given instead of the declination. The Right Ascension is measured from ♈ in the direction of the sun's motion. Right Ascension is usually written in the abbreviated form R.A., and North Polar Distance, N.P.D.

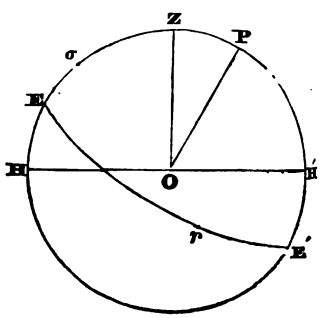
23. The position of σ may be equally well determined by means of the arcs ♈ L and $L\sigma$, which are called respectively its *longitude* and *latitude*.

24. By measuring R.A. and longitude from ♈, we appear to assume that ♈ is a fixed point. This is not accurately true, as we shall see hereafter; its motion however is sufficiently slow to allow us, in general, to conceive it to be fixed; in making observations the motion is allowed for.

ON THE MODE OF MAKING OBSERVATIONS OF THE HEAVENLY BODIES.

25. We propose to explain the mode in which the Right Ascension and Declination of a star become matters of observation, and to describe the principal instruments by means of which the observations are made.

Let Z be the zenith of the place of observation, O the centre of the celestial sphere, E ♈ E' the equator, P its pole, HZP the meridian of the place.

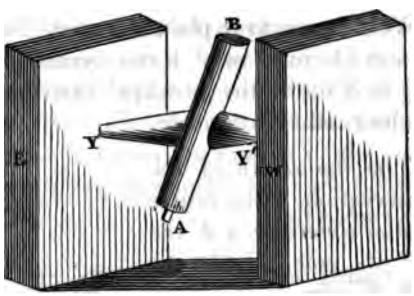


Let σ be the star or other heavenly body, the R.A. and declination, or (N.P.D.,) of which we wish to determine, and suppose it to be on the meridian when we make our observations. Then by means of an instrument called the Mural Circle, soon to be described, we can observe with great accuracy $Z\sigma$, the zenith distance of σ, and this added to ZP, the colatitude of the place of observation, which may also be supposed

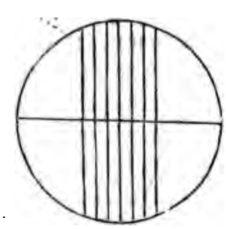
known, will be the N.P.D. of σ. Again, since the whole heavens turn about OP uniformly in 24 hours, called *sidereal* hours, if we note the time of ♈ passing the meridian and that of σ, the difference between these times will measure the arc ♈E, or the R.A. of σ; for instance, suppose the difference of time was 1 hour, then the R.A. would be $\frac{360^0}{24}$, or 18^0. The method adopted in practice is to have a clock which indicates $0^h\ 0^m$ when ♈ is on the meridian, and then the time of a star's passing the meridian, or the time of its *transit*, converted into degrees at the rate of 18^0 to 1 hour, will be the R.A. of the star.

Thus the determination of the R.A. and N.P.D. of a heavenly body is reduced to that of the time of passing the meridian, and the meridian zenith distance. We shall proceed to describe the Transit Instrument and the Mural Circle, by means of which this is effected.

26. *The Transit Instrument.*



This instrument consists of a common astronomical telescope, fixed firmly to an arm YY', the extremities of which are cylindrical and are supported by two pillars of solid masonry E, W, so placed that the arm YY' points east and west. In the focus of the object-glass are placed a certain number of fixed vertical wires, usually five or seven, and one horizontal wire



passing through the centre of the field, as in the annexed figure.

The telescope moves, as will be seen from the preceding description, in the plane of the meridian, and the mode of observation is as follows: before the object to be observed comes upon the meridian, the transit instrument is set in such a position that the object shall pass as near as possible through the centre of the field of view; this may be done by means of a variety of contrivances, but it requires an approximate value of the zenith distance of the object, which however may be supposed known from previous observations sufficiently nearly for the purpose, in the case of all bodies of which it is necessary to observe the transit. The time of transit is the moment at which the object crosses the middle vertical wire, supposing that wire to be in perfect adjustment; but to avoid the error of imperfect adjustment, and also to diminish as far as possible the errors of observation, it is usual to note the time of transit across each of the vertical wires, and take the mean of these observed times as the true time of transit.

The observation requires the assistance of a sidereal clock which beats seconds in a very distinct manner; the observer before looking into the telescope takes notice of the time indicated by the clock, and by counting the beats of the pendulum knows the time which elapses afterwards; having a book and pencil in his hand, he thus notes with great accuracy the time of transit over each wire.

27. The transit instrument is in perfect adjustment when the *line of collimation*, that is, the line joining the intersection of the horizontal and middle vertical wire with the centre of the object-glass, moves in the plane of the meridian, the instrument being turned about its horizontal axis.

In order therefore that the adjustment may be perfect, (1) the line of collimation must be accurately perpendicular to the geometrical line in the transverse axis about which the instrument turns; (2) this geometrical axis must be accurately horizontal; (3) the same axis must point accurately east and west.

In practice the transit instrument is seldom or never in perfect adjustment, and even if it could be made perfect at any given time, so small a cause is sufficient to sensibly disarrange it that the adjustment would soon cease to be perfect. The three errors therefore corresponding to the three adjustments above-mentioned, and which are known as the errors of *collimation*, *level*, and *deviation*, respectively, are ascertained and correction made in the observed time of transit.

28. *The Sidereal Clock.*

The clock when properly adjusted ought to indicate $0^h\ 0^m\ 0^s$ when the first point of Aries is on the meridian, and should indicate the lapse of 24 *sidereal* hours during the interval of two successive transits of that point. The time of the transit of a star according to the sidereal clock will be the star's R.A. in *time*.

The clock is corrected by observing the time of transit of a star, the R.A. of which is accurately known, and comparing the known R.A. with that indicated by the clock.

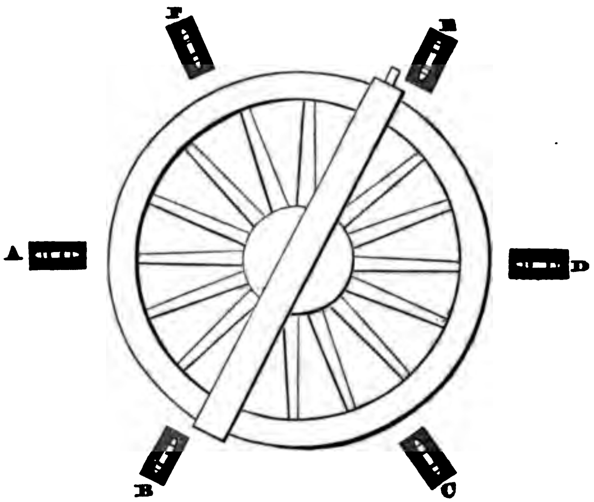
By observing two transits of the same star, or the transit of two known stars, we can ascertain the clock's *rate*, that is, the rate at which it is gaining or losing.

And by observing three transits, we can ascertain whether its rate is regular.

29. *On the Mural Circle.*

This instrument consists of an astronomical telescope firmly clamped to a flat circular rim, which moves about an axis passing into a strong vertical wall to which the back of the instrument is applied. The graduation is made on the side of the rim perpendicular to the wall, and the reading off is effected by means of microscopes *A*, *B*, *C*, *D*, *E*, *F*. In the focus of the object-glass are two fixed wires, one vertical and one horizontal, and besides these there is a moveable

horizontal *micrometer* wire, the nature of which will require some explanation.



A micrometer wire is one which is moveable by means of a screw within the observer's reach, the head of which is graduated and so contrived as to indicate the distance through which the wire is moved. Suppose, for instance, in the above case, when the micrometer wire coincides with the fixed horizontal wire, the pointer on the screw-head indicates $0^0\ 0'\ 0''$, then if when the wire is made to coincide with any given object in the field of view the reading of the screw-head is n'' we shall know that the distance between the image and the fixed horizontal wire subtends an angle of n'' at the centre of the object-glass, in other words, that the angular distance between the object and the centre of the field of view as measured by an arc of a great circle on the celestial sphere is n''.

30. Since the purpose of this instrument is to observe the zenith distance of heavenly bodies as they pass the meridian, the line of collimation ought to move in the plane of the meridian. But it is not difficult to see that the error occa-

sioned by a slight deviation of the line of collimation from that plane will not be so important as in the case of a transit instrument: the error of principal importance is this; when the instrument is pointed to the zenith, one of the microscopes ought to indicate accurately 0″, and this will not generally be the case, the error in the reading may be called the error of collimation in altitude, and may either be calculated and allowed for, or may be got rid of entirely by a method of observation which we shall presently describe. Properly speaking, the error which we have called that of collimation in altitude consists of two, one arising from the fact that if the line of collimation were in perfect adjustment there would in general be an *index error*, and the other from the usual want of accurate adjustment of the line of collimation; these two however are necessarily joined together, and may be conveniently spoken of as one error.

31. The reading off is made with six microscopes instead of one, for the sake of greater accuracy, especially in these respects; we diminish the probability of error from defective graduation; but still more we avoid any error arising from false centering, i. e. from the centre about which the circle turns not coinciding with the actual centre of the graduated rim, for it will be easily seen that if on this account any microscope gives a reading in excess, the opposite one will give a reading as much in defect, and thus the mean of the two will be correct. The mode of reading by means of a microscope is simply this: the part of the graduated rim, the image of which coincides with the centre of the field of the microscope, is that the position of which we desire to determine; but this will in general not happen to coincide with a line of graduation, we have therefore to determine the distance of the centre of the field from the nearest line of graduation; this is done by means of a micrometer wire, which when it coincides with the centre of the field gives a reading on the screw-head of 0″, and hence if we turn the screw-head until the micrometer wire coincides with the image of the nearest line of graduation, the reading on the screw-head will be the quantity to be added to the reading of the circle, correspond-

ing to the nearest line in question. The first reading is thus made with the naked eye, and the microscope then furnishes a correction to this reading.

32. The mode of observation with the mural circle is as follows.

Before the star, or other heavenly body to be observed, crosses the meridian, let the telescope be directed downwards towards a trough of mercury, in such a manner that the image of the star seen by reflexion at the surface of the mercury may pass through the field of view: this can be done, because the zenith distance is supposed to be known approximately. Let the circle be clamped in this position, and the microscopes read.

When the star, as seen by reflexion, comes into the field of view, let it be bisected by the moveable micrometer wire, a short time before it reaches the middle of the field.

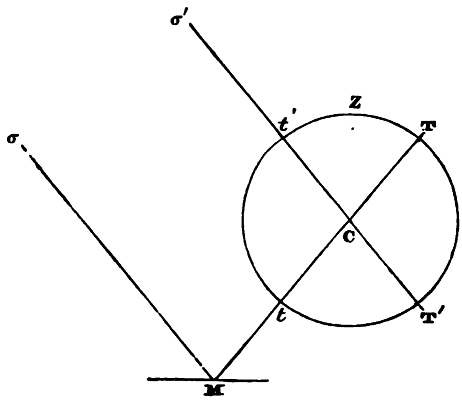
Let the circle be now unclamped, and turned rapidly round until the star is visible by direct vision, and by means of a screw provided for the purpose let the circle be slightly moved until the star is bisected by the fixed horizontal wire. This may be effected by a skilful observer soon after the star has passed the middle of the field.

The microscopes may now be read, and the mean of the six readings will be the result of the observation by direct vision. The micrometer may also be read, and this reading added to the mean of the readings of the six microscopes, as examined before the observation, will give the result of the observation by reflexion.

The mean of the two observations will give the true altitude or zenith distance of the star, free from error of collimation in altitude. That this will be so may be seen as follows.

33. Let TtM be the position of the line of collimation, when the star σ is seen by reflexion. Then if Z be the zenith point, the reading ought to be Zt, but in consequence

of the collimation error let it be $Zt + C$. Also let $T't'$ be the position of the line of collimation, when the star is seen



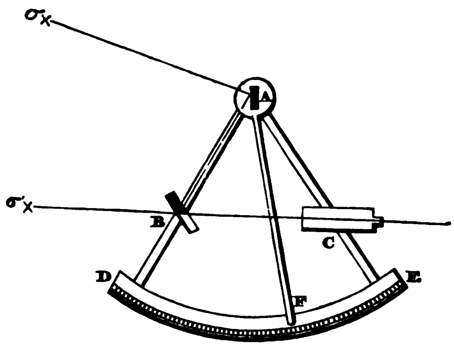
by direct vision, and the reading will be $Zt' + C$; hence the mean of the readings $= \dfrac{Zt - Zt'}{2} = \dfrac{tt'}{2} =$ the true altitude of the star.

34. The Transit Instrument and the Mural Circle are the instruments of daily use in observatories, and we recommend the student, if it should be within his power, to visit some observatory for the purpose of inspection of the instruments. We shall conclude this part of the subject, by describing Hadley's Sextant, an instrument of great importance, on account of its applicability to nautical purposes; the instruments already described, and all others which require fixed supports, are useless at sea on account of the constant motion of the vessel.

35. *Hadley's Sextant.*

The principle on which this instrument is constructed, depends upon a proposition proved in Optics, (Art. 27, p. 363), viz. that when a ray of light is reflected at two mirrors, the angle of deviation is equal to twice the angle between the mirrors.

The figure represents Hadley's Sextant; *DE* is a portion of a graduated rim, the graduation, as the name imports,



extending usually to about one sixth of a circumference, but sometimes to more. *AF* is a moveable radius of the circle, which carries with it a small mirror *A* of silvered glass. *B* is a piece of glass silvered over half its surface, and is so fixed that when the moveable radius points to the zero of the graduation, the surfaces of the mirrors *A* and *B* are parallel. *C* is a small telescope attached to the instrument, and so arranged that its axis passes through the line of division between the silvered and unsilvered parts of the glass *B*.

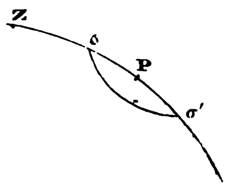
To find the angle subtended by the line joining two objects, σ, σ', let the instrument be held in such a position, and the moveable radius so adapted, that the image of σ as seen by reflexion at the two mirrors, coincides with that of σ', as seen by direct vision; then the angle between the two objects will be twice the angle between the mirrors, or twice the arc between *F* and the zero-point, since when the mirrors were parallel *F* pointed to zero. Consequently, if we graduate the arc *DE* in such a manner as to make one of its divisions correspond to 2^0, the reading given by *F* will be the angle required.

By means of this instrument the angular distance between two objects, or the altitude of a heavenly body above the horizon, may be ascertained with sufficient exactness for nautical purposes.

ON METHODS OF FINDING THE LATITUDE OF A PLACE.

36. In our explanation of the mode of finding the N.P.D. of a heavenly body, we have assumed the latitude of the place of observation to be known; we proceed to shew how it may be ascertained.

Let σ σ' be the small circle described by a circumpolar star, round the pole P, in consequence of the diurnal revolution of the heavens; Z the zenith of the place of observation, $Z\sigma P\sigma'$ its meridian.



When the star is on the meridian at its upper transit, let its zenith distance $Z\sigma$ be observed; and when it is again on the meridian at its lower transit, let its zenith distance $Z\sigma'$ be observed.

Then $$ZP = \frac{Z\sigma' + Z\sigma}{2} = \text{the colatitude};$$

hence the latitude is known.

The preceding method is applicable, only when we have a fixed instrument in the plane of the meridian, as in an Observatory. Methods applicable to observations made at sea require mathematical calculations, into which it is not the purpose of this treatise to enter.

ON THE FIXED STARS.

37. If we look at the heavens night after night, we shall easily conclude that the greater part of the stars preserve, at least approximately, the same relative positions, and if we determine their R.A. and N.P.D., we find them nearly constant; such stars are called *fixed* stars, a name which distinguishes them from the *planetary* bodies, of which we shall afterwards have to speak more particularly. The modes of

determining the R.A. and N.P.D. of a star, which we have described, will require considerable correction, but, supposing that we have the means of determining accurately these elements, we can make a catalogue of the fixed stars. Such catalogues have been made, and are of the utmost use in practical Astronomy; the place of a star becomes known accurately, only by a long series of observations, and when the R.A. and N.P.D. of a star have thus been satisfactorily established, it is called a *known* star. These fixed stars are at an enormous distance from the earth, as is at once concluded from this fact, that the most powerful telescope does not exhibit any sensible disk; but of the greatness of their distance we shall presently have a more definite notion, when we come to speak of parallax. The stars are divided into groups called constellations; this however is not done upon any good and convenient system, but in a manner apparently quite arbitrary, and certainly very fanciful. If we refer the places of the stars, and those of the sun, moon, and planets, to the surface of the celestial sphere, the stars will be fixed in position, and the sun moon and planets will move amongst them; hence we may speak of the motion of the sun among the fixed stars, and the signs of the zodiac are in fact the names of twelve constellations, which at the time the names were given were coincident in position with the signs bearing the same names, but now, for reasons to be hereafter assigned, occupy other positions.

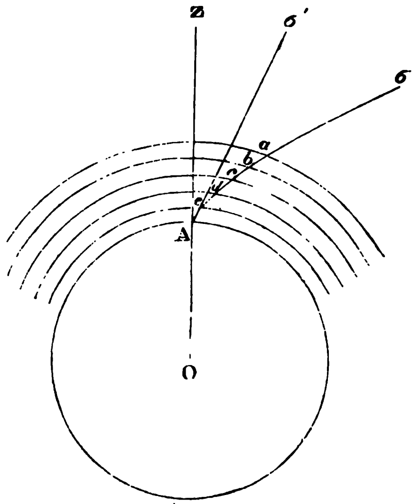
The latitude and longitude of a star may be deduced from its right ascension and declination, and stars may be catalogued accordingly, but this method is not so convenient as that of registering their right ascensions and declinations.

ON THE CORRECTIONS OF ASTRONOMICAL OBSERVATIONS.

38. We have described the mode of determining by observation the R.A. and N.P.D. of any heavenly body, but the observations so made require several important corrections, which we now proceed to explain.

ON REFRACTION.

39. The first correction is due to the deviation caused in a ray of light by its passage through the atmosphere.



Let O be the centre of the earth, A the place of observation, Z its zenith. And conceive the atmosphere to be formed of concentric strata of air, diminishing in density as they are further from the earth's surface.

Then a ray of light in passing through the atmosphere will, on entering each stratum, be bent towards the normal, (Optics, Art. 7, page 351), and consequently the ray of light by which the object σ is seen by an observer at A will be of the kind represented by the line $\sigma abcdeA$. In the limit, when we consider the strata to be indefinitely thin, the path of the ray through the atmosphere will be a continuous curve, and the direction in which the object will be seen will be the tangent to the curve at the point nearest the observer's eye.

The atmosphere will be exactly similar on opposite sides of the vertical plane through $ZA\sigma$, and therefore the ray of light will not be refracted out of that plane. Hence a heavenly body appears to be raised in a vertical plane above its true position, and the zenith distance given by observation will be

28—2

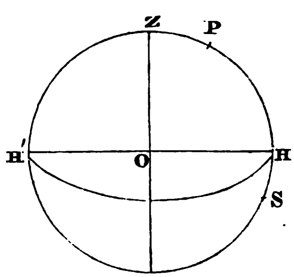
too great; the amount of correction to be applied depends upon calculations into which we shall not enter.

40. *On the phenomenon of Twilight.*

The consideration of the effect of refraction upon the directions of rays of light passing through the atmosphere, renders this a convenient place to explain the phenomenon of twilight. We have shewn in the preceding article, that the heavenly bodies are apparently raised in a vertical plane by the refraction of the atmosphere, and it follows that they become visible on the earth's surface when they are in fact somewhat below the observer's horizon. But before it becomes visible, the rays of light from the sun below the horizon illuminate the atmosphere, which to some extent reflects and scatters the rays in all directions, and the result is a faint light which precedes the rising of the sun and follows its setting, and which we call *twilight*. Twilight begins and terminates when the sun is about 18^0 below the horizon; but its duration manifestly varies with the latitude, for the time which is required for the sun to rise through 18^0 vertically depends upon the inclination of its diurnal path to the horizon of the place, and is less as this inclination is greater, that is, as the place is nearer to the equator. To take an extreme case, suppose the place to be on the equator, and for simplicity's sake, suppose the sun to be in ♈, then its diurnal course will be a vertical great circle, and the duration of morning or evening twilight will be time taken to describe 18^0 or not much more than an hour.

In more northern latitudes the twilight may endure all night; it is not difficult to determine the conditions under which this will take place.

Let Z be the zenith of the place, $H'H$ the horizon, P the pole of the equator, S the sun at midnight. Then the condition of twilight lasting all night is, that the greatest depression of the sun below the horizon shall not be more than 18^0, or HS not greater than 18^0.



Let l be the latitude of the place $= HP$,

δ the sun's declination $= 90^0 - PS$:

then $HS = PS - HP = 90^0 - \delta - l$;

$\therefore$ we must have

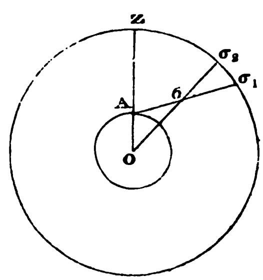
$$90^0 - \delta - l \text{ not} > 18^0,$$

or $$\delta + l \text{ not} < 72^0.$$

For instance, the latitude of London is $51\frac{1}{2}^0$, therefore twilight endures all night so long as the sun's declination is not less than $20\frac{1}{2}^0$, that is, from the latter end of May to the latter end of July.

ON PARALLAX.

41. This second correction is rendered necessary by the fact of our observations being made at the surface of the earth, and not at the centre. If the position of a heavenly body were registered according to its apparent position as seen from a certain place, the position so determined would not agree with that determined by another observer at a different place on the earth's surface; a correction is therefore applied to observations, such as shall reduce them to what they would have been if made from the earth's centre. This we proceed to explain more particularly.



Let O be the earth's centre, A the position of an observer on its surface, Z the zenith; and let σ be some heavenly body, which an observer at A will refer to the point σ_1 on a sphere described with centre O, and an observer at O will refer to σ_2. Hence the zenith distance of σ as observed from A will have to be diminished by the quantity $\sigma_1 \sigma_2$ in order to reduce

it to what it would have been if observed from O. Parallax, it is evident, takes place in a vertical plane, and depresses a heavenly body, which is exactly the opposite effect to that of refraction; in speaking, however, of parallax as compared with refraction, it is to be carefully borne in mind that they are corrections in very different meanings of the word, for, in consequence of refraction, the object is not actually in the position in which it seems to be, whereas the correction for parallax is merely a reduction of the observations made at one place to what they would have been if made at another.

42. In the figure let $p = A\sigma O = \sigma_1\sigma_2$ nearly, $ZA\sigma = z$, then, from the triangle $A\sigma O$,

$$\sin p = \frac{AO}{O\sigma} \sin z;$$

or since p is very small, if we use the circular measure,

$$p = \frac{AO}{O\sigma} \sin z. \quad \text{(See page 270, note).}$$

This quantity p, since it varies with the altitude of the body, is called the *diurnal* parallax; its greatest value is when the body is in the horizon or $z = 90^0$, in which case (if we call the value P),

$$P = \frac{AO}{O\sigma}, \text{ and } p = P \sin z;$$

P is called the *horizontal* parallax.

43. It is found that some of the heavenly bodies, namely, the fixed stars, have no sensible horizontal parallax, but for the sun, moon, and the planets, (of which we shall hereafter speak more particularly,) the value is sensible, and the determination of its value in each case becomes a matter of great moment; for it will be seen, that the knowledge of the horizontal parallax of a heavenly body informs us at once of the

distance of that body from the earth's centre in terms of the earth's radius, the magnitude of which we shall presently shew how to find. In fact, we have by the last article

$$O\sigma = \frac{AO}{P};$$

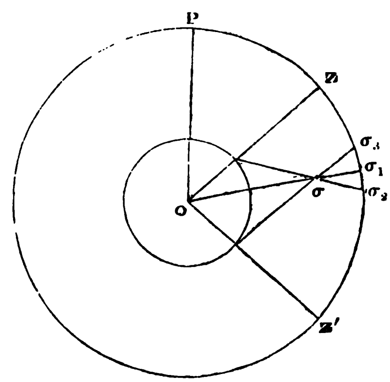
and if R be the earth's radius, x the distance of the heavenly body, and the horizontal parallax be given in seconds,

$$x = \frac{180 \times 60 \times 60}{P \times 3.14159} R.$$

(See Trig. Art. 54, page 125.)

The value of the sun's horizontal parallax is about 8.75″, that of the moon's 57′ 4″.

44. *To find the horizontal parallax of a heavenly body by observation.*



Let Z Z' be the zeniths of two places on the same meridian PZZ': and let σ be a heavenly body upon the meridian, which is referred by observers at the two places to the points σ_2 σ_3 on the celestial sphere respectively, its true position as seen from the centre of the earth being σ_1. Then if P be the horizontal parallax of σ, we have

$$\sigma_1\,\sigma_2 = P \sin Z\,\sigma_2,$$

$$\sigma_1\,\sigma_3 = P \sin Z'\sigma_3;$$

also
$$Z\sigma_2 - \sigma_1\,\sigma_2 + Z'\sigma_3 - \sigma_1\,\sigma_3 = ZZ',$$

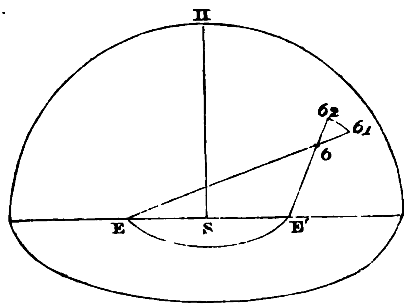
$$\therefore\quad P = \frac{Z\sigma_2 + Z'\sigma_3 - ZZ'}{\sin Z\sigma_2 + \sin Z'\sigma_3}.$$

In this equation $Z\sigma_2$ $Z'\sigma_3$ are known by observation, and ZZ' is the difference of the latitudes of the places, (or the ***sum***

if the places are on opposite sides of the equator, as in the figure;) hence P is completely determined.

This method is applicable to the moon, the parallax of which is considerable; but the sun's parallax being a very minute quantity, can only be satisfactorily determined by the help of very particular phenomena, a description of which we shall not here enter upon.

45. The diurnal parallax of the fixed stars is found to be wholly insensible: but there is another kind of parallax due to the change of position of the earth in her orbit, which is called *annual* parallax, and it becomes a question how far this will affect the apparent position of a fixed star.



Let S be the sun, E, E' two positions of the earth on opposite sides of it; σ a star which an observer on the earth at E refers to the point σ_1, and an observer on the earth at E' to the point σ_2 on the celestial sphere; then $E\sigma E'$ is the greatest variation in the position of σ caused by parallax, and as the distance EE' is about 190,000,000 miles, we should expect that this angle would be sensible: but so enormous is the distance of the fixed stars that ordinary observations detect no annual parallax. By observations of extreme delicacy, Bessel, the eminent astronomer of Königsberg, believes that he has detected parallax in one star (61 Cycni); this is a double star, that is, when seen through a good instrument it appears

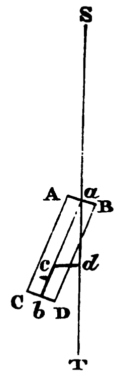
to consist of two distinct bodies revolving about each other, or rather about the centre of gravity of the two, and this was one reason for its selection for the purpose of attempting to detect parallax, because the distance of the point halfway between the two stars from any given point admits of more exact determination than that of the distance of a star itself: another reason for choosing this star was, that it has a larger amount of *proper motion*, that is to say, its place in the heavens instead of being absolutely fixed, is, after every correction has been made, found to vary regularly from year to year, and to whatever cause this may be due, a large amount of proper motion would seem almost certainly to indicate proportionate proximity to our system. (See page 448.) The mode of observation was to determine the distance of the point midway between the stars from two other small stars, which it was considered might be supposed to be free from sensible parallax, and by continuing the observations for a year, it was found that after every allowance had been made, there appeared to be a change of position due to annual parallax. The result is, that the star 61 Cycni appears to have an annual parallax of about $0''.314$, which gives its distance from the sun to be 657,700 times that of the earth from the sun, or 62,484,500,000,000 miles: light takes rather more than 10 years to traverse this space; yet this is probably one of the nearer of the fixed stars. The period of revolution of the component stars about their centre of gravity is about 540 years; hence, concluding their distance from each other from the preceding calculated distance from the earth and their apparent angular distance from each other, we can deduce the sum of their masses, which proves to be about half that of the sun.

ON ABERRATION.

46. The third correction to be applied to observations is due to the fact of light travelling with a finite velocity; the manner in which this circumstance affects the apparent position of a star may be explained as follows.

Suppose a particle to move with a uniform velocity from S towards T, and suppose it is required to hold a tube $ABCD$, which is also in uniform motion in a direction perpendicular to ST, in such a position that the particle shall move along its axis. Let a, a point in the axis, be the place of the particle at any given time; let c be any other point; draw cd perpendicular to ST, then if the inclination of the axis of the tube to ST be such that

$cd : ad ::$ velocity of tube : velocity of particle,



it is manifest that when the particle has arrived at d, c will also have arrived there, and the particle will always be on the axis of the tube, as was required.

Let θ be the angle of inclination of the tube, v the velocity of the tube, V of the particle, then we must have

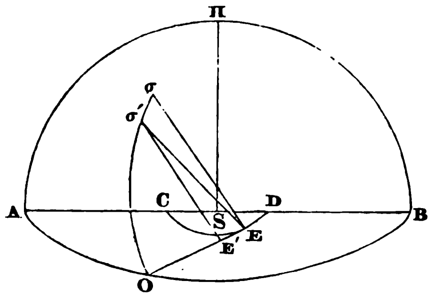
$$\tan\theta = \frac{v}{V}.$$

Now the case we have supposed is nearly analogous to that of a heavenly body viewed through a telescope; for the telescope is in motion in consequence of the motion of the point of the earth's surface on which it is fixed, and in order that a star may be visible through it, we must hold it in such a direction that a ray of light coming from the star may pass down its axis; hence we conclude from what precedes, that the direction of the axis of the telescope does not coincide with the direction in which light proceeds from the star, but is inclined at angle (θ) determined by the equation

$$\tan\theta = \frac{v}{V},$$

where v is the velocity of the earth in its orbit, and V the velocity of light. If then v be comparable with V, there will be a consequent error in our observations, and this error is said to be due to *aberration*. We shall shew how it can be ascertained by observation that the error is appreciable, but first we must explain more particularly the effect of aberration in altering a star's apparent place on the celestial sphere.

47. Let AOB be the ecliptic, Π its pole, S the sun, E the earth in its orbit CED, σ a fixed star. Draw EO a



tangent to the earth's orbit, then EO is the direction of the earth's motion, and σE that of light from the star, hence aberration takes place in the plane σEO; join σO by an arc of a great circle, and let σ' be the apparent place of the star. Join $\sigma' E$, and draw $\sigma' E'$ parallel to σE; then, by what has been said, we must have

$$EE' : \sigma E\,(=\sigma' E) :: \text{velocity of earth} : \text{velocity of light};$$

$$\therefore \sin\sigma\sigma' = \sin\sigma E\sigma' = \sin E\sigma' E = \frac{EE'}{\sigma' E}\sin\sigma' EO,$$

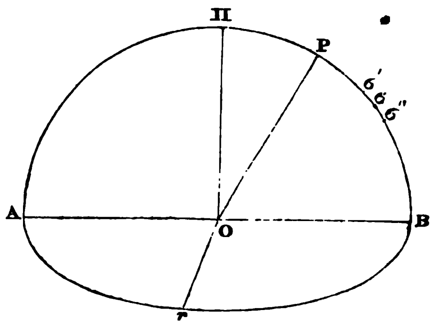
$$\text{or } \sigma\sigma' = \frac{\text{velocity of earth}}{\text{velocity of light}}\sin\sigma O, \text{ nearly.}$$

The angle σO is called the *earth's way.*

The velocity of the earth in its orbit is so much greater than that of any point in its surface due to rotation about the axis, that we may confine our attention to the former as producing the error of aberration.

It will be easily seen, that the effect of aberration is to make the apparent place of a star describe a small curve about its real place in the course of a year.

48. Let us now examine a particular case: suppose σ to be on the solstitial colure $A\Pi PB$, and the earth to be in ♈, as at the autumnal equinox. Then the earth is moving towards



A parallel to the solstitial colure, and consequently the star σ is displaced by aberration into the position σ', if

$$\sigma\sigma' = a \sin \sigma A = a \sin \sigma B; \ \left(\text{where } a = \frac{\text{velocity of earth}}{\text{velocity of light}}\right).$$

Again, at the vernal equinox the apparent place of the star will be σ'', if

$$\sigma\sigma'' = a \sin \sigma B = \sigma\sigma'.$$

Hence we are furnished with the means of determining the coefficient of aberration (a) without previous knowledge of the velocity of light. For let dd' be the N.P.D. of the star at the autumnal and vernal equinox respectively, then

$$d' - d = \sigma'\sigma'' = 2\sigma\sigma' = 2a \sin \sigma B = 2a \cos (d + \omega),$$

$$(\omega \text{ being} = \Pi P, \text{ the obliquity of the ecliptic});$$

$$\therefore \ a = \frac{d' - d}{2 \cos (d + \omega)}.$$

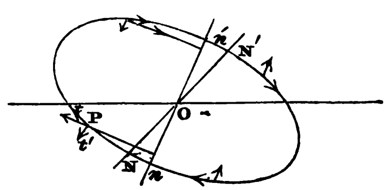
49. Hence, also, we are able to determine the velocity of light, for it may be deduced at once from the value of a, if the velocity of the earth in its orbit be known; but this is determined by the distance of the earth from the sun, which is known from the value of the sun's parallax. Thus the phenomenon of aberration enables us to determine the velocity of light, and the coincidence of the value thus obtained with that deduced by a perfectly independent method, which will be

mentioned in the sequel, leaves little doubt concerning the truth of the principle of both investigations.

ON PRECESSION AND NUTATION.

50. The preceding articles contain an account of the three astronomical corrections which are necessary to be applied to any observation; but besides these, another correction, though of a different kind, must be applied to the R.A. and N.P.D. of a star, as determined by observation, in order to make them coincide with those registered in a catalogue of stars. To understand this, it is only necessary to observe, that the R.A. and N.P.D. of a star, when considered as fixing its position, necessarily assume the pole of the equator and the point ♈ to be fixed with regard to the ecliptic, which they are not. We have already alluded in a cursory manner to this fact; we shall now attempt an exposition of its physical cause.

We shall hereafter shew how to prove by observation a fact which has been already mentioned, and which we shall here assume, namely, that the figure of the earth is not accurately spherical, but that of an oblate spheroid or surface generated by the revolution of an ellipse about its minor axis.



51. Let us first consider the motion of a hoop, which is revolving uniformly in its own plane about its centre O, supposed stationary: and let NON' be the line of intersection of the plane of the hoop with a certain fixed plane through O. Now suppose each particle of the hoop to receive a sudden small impulse perpendicular to its plane, and tending to bring it to coincide with the fixed plane before-mentioned; let us see what the effect will be. Consider any particle P which is moving in the direction Pt, and receives a sudden

impulse in the direction Pt', then the resultant motion of P will be in a direction intermediate to Pt and Pt', and producing this direction to meet the fixed plane, we see that the line NN' will have *regreded* by the impulse into the position $n\,n'$. The angular velocity of the hoop will not be affected, because the impulse on each particle is perpendicular to the motion of the particle. Moreover the inclination of the plane of the hoop to the fixed plane will not be affected, for whatever tendency there is in the impulse on a particle P, which is moving *from* the plane, to diminish the inclination, there is exactly the same tendency in the impulse on a particle P', which is moving *towards* the plane, to increase it. If instead of a single impulse, we suppose continuous forces acting in the manner described, namely, always tending to bring the plane of the hoop into coincidence with the fixed plane, the result will be, that the angular velocity and inclination will be unaltered, and the line of intersection of the two planes will regrede on the fixed plane, that is, will move in the opposite direction to that of the hoop's revolution.

52. The motion which we have described will also be that of a body shaped like the earth, which may in fact be conceived of as a sphere surrounded by hoops of matter, producing the protuberance at its equator. And the attractive force of the sun has always a tendency to make the plane of the equator coincide with that of the ecliptic, as may be seen with a very little consideration, since it tends to bring all particles into the plane in which it is itself moving. Hence the result of the sun's attraction on the protuberant matter, in the mid-regions of the earth's surface, is to make the first point of Aries *regrede*, that is, move in a direction opposite to that of the earth's rotation, or of the sun's motion, without affecting the length of the day, or the obliquity of the ecliptic. This regression of ♈ gives rise to what is called the *precession of the equinoxes;* for since ♈ moves in the opposite direction to the sun, it goes as it were to meet the sun, and consequently the time of arriving at the equinox will precede the time at which the sun would reach it if it

were stationary. The whole phenomenon of the motion of the first point of Aries is called *Precession;* the motion is about 50″ annually. It is not difficult to see, that in consequence of the regression of the line of intersection of the equator and ecliptic, the pole of the former will describe about that of the latter a small circle of the celestial sphere; in other words, the straight line drawn through the earth's centre, perpendicular to the plane of the equator, will describe about the line perpendicular to the plane of the ecliptic, a cone having for its semi-vertical angle the obliquity of the ecliptic.

53. The results of this precessional motion are remarkable: it will be easily seen that the fixed stars will change their position, with respect to the equator and its pole, and the change of their position will be represented by supposing the whole celestial sphere to revolve with a slow angular motion from west to east, about an axis passing through the pole of the ecliptic; the time of this revolution is about 26,000 years. As an illustration it may be observed, that it is only an accident of the present age, that we have what is called a pole star; the star which is now so near the pole, will, after a series of years, leave that position, and only return to it after a complete revolution of the heavens.

It may also be observed, that when the names, Aries, Taurus, &c., were given to the signs of the zodiac, the beginnings of those signs were found in constellations bearing those names; but now the *signs* are far distant from the *constellations* which have respectively the same names.

54. It is not difficult to see, that the action of the sun on the earth is not exactly the same for all periods throughout the year; the mean effect on the pole of the equator will be such as has been described, but to represent the facts more accurately, we must suppose the pole to describe about the pole of the ecliptic, not a small circle, but a tortuous curve, sometimes approaching the pole of the ecliptic, sometimes receding from it, this curve, however, never differing

much from a circle. This irregularity of the motion of the pole is called its *nutation*.

55. There will be, in like manner, a Precession and Nutation of the pole of the equator due to the action of the moon, but these are not nearly so important as those due to solar action, and will not be further considered.

ON THE PROPER MOTION OF THE FIXED STARS.

56. If the R.A. and N.P.D. of any fixed star be found by observation, and all the corrections which we have described be applied, it will nevertheless be found that the position of the star is not actually fixed. The change of position is very small and regular, and for many stars is a known quantity. No physical cause has been certainly assigned, but the variation is attributed to a proper motion of the stars themselves. And reflection will shew that such a proper motion might have been expected *à priori;* for supposing the law of attraction which holds in our own system to extend, as most probably it does, to the regions of the fixed stars, then the mutual attractions of the stars must prevent them from being actually fixed in space. Nevertheless, in consequence of their enormous distance, the proper motion of a star will produce an extremely small apparent change of position to an observer on the earth's surface, and therefore we still use the name *fixed* stars, although believing the bodies so called to be really in motion.

ON THE MODE OF DETERMINING THE EARTH'S FIGURE AND MAGNITUDE.

57. We shall first explain how the radius of the earth may be found, supposing its form to be spherical.

We have already explained (Art. 36), how to determine the latitude of a place; now let two places be found on the same meridian, the difference between the latitudes of which is

1^0, and let the distance between these two places be measured as accurately as possible, and be found to be D. Let R be the earth's radius, then its circumference is $2\pi R$, where $\pi = 3{\cdot}14159......$ (see Trig. Art. 51, page 122); and since angles are proportional in the same circle to the subtending arcs, we have

$$\frac{2\pi R}{D} = \frac{360^0}{1^0};$$

$$\therefore\ R = \frac{180}{\pi} D,$$

which equation determines R: from such measurements it is found that $R = 4000$ miles nearly.

58. But it is found that when the distance between two places, the latitude of which differs by 1^0, is measured in different parts of the same meridian, the results do not agree. It is found that the length of a degree (as it is called) increases with the latitude, *i.e.* is least at the equator and increases as we approach the poles; in other words, near the equator it requires a smaller change of our position than near the poles to produce the same amount of change in the position of the zenith: now it is manifest, that the more the surface of the earth is curved the more rapid will be the change of zenith to a person walking over it, hence we conclude that the curvature of the earth is greatest at the equator, and diminishes towards the poles. Hence the equatoreal radius must be greater than the polar radius, and the earth may be conceived of in general as a globe slightly flattened at the poles. Accurate observation shews that the rate, at which the curvature of the earth's surface changes, agrees with the hypothesis of its form from being that of an oblate spheroid, as has already been assumed.

ON THE SOLAR SYSTEM.

59. We shall now proceed to give a more accurate account of the sun and the bodies revolving about it, of which the earth is one.

The principal planets are named as follows, the order being that of their distances from the Sun, Neptune*, Uranus, Saturn, Jupiter, Mars, the Earth, Venus, Mercury. Besides which there are five very small planets, called *asteroids*, and bearing the names, Ceres, Pallas, Vesta, Juno and Astræa; these are invisible to the naked eye, and in point of distance from the sun are intermediate to Mars and Jupiter. Two of these, Neptune and Astræa, have been very recently added to the list, and even now others may exist which have not been recognised, for among the multitude of telescopic stars few, comparatively speaking, have been sufficiently observed to enable us to decide upon the constancy of their position.

60. The paths of the planets, with the exception of the asteroids, are in planes making a small angle with the plane of the ecliptic, and their motions are governed by the three following remarkable laws, called from their discoverer Kepler's Laws.

I. *The planets move in ellipses, each having the sun's centre in one of its foci.*

II. *The areas swept out by each planet about the sun are, in the same orbit, proportional to the time of describing them.*

III. *The squares of the periodic times are proportional to the cubes of the major axes.*

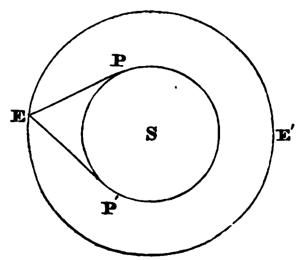
It will be easily seen that these three laws follow at once from the hypothesis of a force, varying inversely as the square of the distance, resident in the sun's centre. (See *Newton*, Props. I. XI. and XV.) Speaking accurately, the centre of the sun is not a fixed point, the motion taking place in fact about the centre of gravity of the whole system; but on account of the enormous mass of the sun, as compared with any one or with

* The title of Neptune has been applied to the new planet, the discovery of which has been made during the passage of the present work through the press, as being that which appears to be most likely to be agreed upon as the permanent name. The discovery forms one of the most remarkable epochs in the history of Astronomy.

all of the planets, the centre of gravity is very near the centre of the sun, and may generally be conceived of as coincident with it. Moreover, no one of Kepler's laws is quite rigidly true, because not only does the sun attract each of the planets, but the planets mutually attract each other, and produce slight perturbations or deviations from laws which would hold strictly for any one undisturbed planet; it may be as well to observe here, that with respect to such perturbations theory and observation are in complete accordance.

61. The planets which are at a greater distance from the sun than the earth, are called *superior* planets, those which are at a smaller distance, *inferior*. It will be anticipated that the phenomena exhibited by these two classes to an observer on the earth's surface will be in many respects different.

Let S be the sun's centre, EE' the earth's orbit, PP' the orbit of an inferior planet, supposed for simplicity's sake to be in the plane of the ecliptic. Then it is evident that the planet's *elongation* from the sun, that is, the angle subtended at the earth by its distance from the sun, can never exceed a certain limit; for from E, the position of the earth, draw EP, EP' tangents to the planet's orbit, then PES, or $P'ES$, will be the maximum angle of elongation, and the elongation will vary between zero and this value; sometimes the planet may even pass between the earth and the sun, and it will then be seen to cross over the sun's face like a dark spot. Now suppose the sun and the planet to revolve round E, then the time of passing the meridian of any given place will not be very different for the sun and the planet; for example, let Venus have her maximum elongation, and suppose her to rise some time before the sun so as to be a *morning* star, then the time by which her rising precedes that of the sun diminishes until at last she is so near the sun that she becomes invisible: afterward she begins to rise after the sun, and therefore to set



after him, and thus becomes an *evening* star. Mercury is seldom sufficiently far from the sun to be visible with the naked eye. There is manifestly no limit to the elongation of the superior planets, and therefore no connection between the times of their rising or setting and that of the sun.

62. The inferior planets present to the earth *phases*, exactly as does the moon; this will be seen to be a necessary result of the manner in which they are illuminated, for since only one hemisphere of the planet, namely that which is turned towards the sun is illuminated, the portion of the illuminated surface visible from the earth must depend upon the relative positions of the sun, earth, and planet. Of the superior planets, Mars, which is the nearest to the earth, presents sometimes a slightly *gibbous* appearance, that is, a phase like that of the moon when near full, but the others have no perceptible changes of phase, because the direction in which light falls upon them from the sun is, on account of their great distance, very nearly the same as that in which they are viewed from the earth.

63. The limits of this treatise will not allow us to enter upon a description of the physical peculiarities of the various planets. We cannot however omit to take notice of the very remarkable object, called Saturn's *ring*. It would be more correct to speak of two rings, for there are in fact two, lying in one plane and separated from each other by a narrow interval, the inner one being separated by a much wider interval from the planet itself. The ring has a rapid rotation in its own plane, the centrifugal force arising from which is doubtless the means of preserving its form and its position.

64. The earth and the superior planets are accompanied by secondary bodies, termed *satellites*, revolving about them in the same manner as they themselves revolve about the sun. Of these satellites or moons, the earth has one, Jupiter four, Saturn seven, Uranus certainly two and perhaps six; whether Neptune is attended by satellites has not been at present ascertained.

For much interesting information on this and other subjects the reader is referred to Sir J. F. W. Herschel's Treatise on Astronomy, being a volume of the Cabinet Cyclopædia; in that work the facts of astronomy are treated with a clearness and liveliness of illustration which cannot fail in the highest degree to interest and instruct.

ON THE MOON.

65. The earth is accompanied (as has been already mentioned) in its course round the sun by a secondary body, which we call the moon, and which revolves round it in the same kind of way as the earth revolves about the sun, and from west to east. The distance of the moon from the earth, as concluded from the value of its horizontal parallax, is about 60 of the earth's radii. The moon revolves round the earth according to the ordinary rules of bodies moving in central orbits, deviating however from the strictness of these rules materially in consequence of the disturbing force of the sun, which is much greater than that of one planet on another, and therefore causes the moon's orbit to deviate more from the elliptical form than is the case with the orbit of a planet. The time of the moon's revolution about the earth is 27 days 7 hours nearly.

66. The most remarkable phenomenon exhibited by the moon is the change of her *phase*, which has been already alluded to in speaking of the phases of the inferior planets, and which is an immediate result of her being illuminated by the sun. It is manifest that when the moon and sun are on opposite sides of the earth, or the moon is in *geocentric opposition*, the bright side of the moon will be turned towards the earth, or it will be *full-moon*; and when the moon is between the earth and sun, or in *geocentric conjunction*, the bright side will be turned away from the earth, or it will be *new-moon*. Between these two extreme cases the moon will present every variety of phase. The time elapsing between two successive full-moons is not that of the moon's revolution, because

the sun has, during the month, advanced in its course; the time will be equal to the time of revolution of the moon, together with that which the moon takes to pass over the space moved through by the sun, or about $29\frac{1}{2}$ days. This is called a *lunation*, or a *synodic period.*

67. It may be noticed that the moon when full, being in exactly the opposite quarter of the heavens from the sun, or in other words, distant from the sun 180^{0} in R.A., comes upon the meridian 12 hours after the sun, i.e. at 12 o'clock at night; hence the moon is brightest at a time when the darkness of the night would otherwise be the most intense.

68. The moon revolves on its own axis in the same direction in which it revolves about the earth, and in the same period; the consequence is that it always presents to the earth the same face. The appearance of the face is extremely ragged and mountainous; the height of some of the mountains has been ascertained, and it thus appears that they are very much greater as compared with the size of the moon, than in the case of terrestrial mountains.

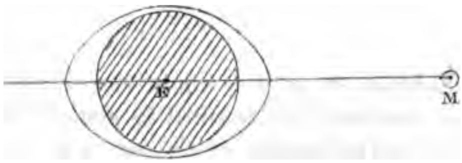
The most obvious office of the moon is that which she performs in giving light to the earth, but an equally important one is that of producing tides, of which we shall presently speak more particularly.

The phenomenon of eclipses also, which the moon is instrumental in producing, will be separately considered.

ON TIDES.

69. The accurate theory of tides is one of the utmost difficulty; but a general explanation, sufficient for many purposes, may be given as follows.

Let E be the earth, M the moon; and suppose the earth to be surrounded by a sea, and the moon to be at rest with respect to it. Then the attraction of the moon on each particle



of water will tend to draw the water away from the earth on the side turned towards it, and thus to make a protuberance of water: there will be a similar protuberance on the opposite side of the earth, because the force which draws any particle of water away from the earth is the difference between the attraction of the moon on that particle and on the earth, and hence the force in question upon a particle on the opposite side of the earth from the moon will tend from the earth; and hence, as we have said, there will be a protuberance of water on the side of the earth which is turned away from the moon, as well as on the other. The accurate form of the surface of the sea is a spheroid, having its larger axis passing through the moon.

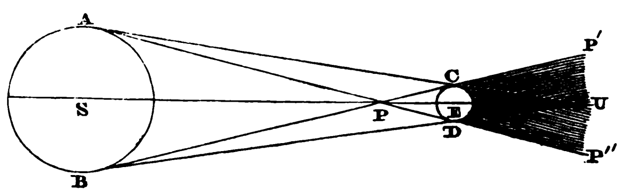
If there were no moon there would be a similar disturbance of the spherical form of the sea by the sun's attraction. In order to determine the result of their combined action, it will be sufficient to observe, that if the sun and moon are on the same side of the earth, or the moon in geocentric conjunction, the action of the two will be combined, and we may suppose the effect to be the sum of those due to their separate actions: but if the sun and moon are on opposite sides, or the moon in geocentric opposition, the effect will be the difference of the two; in other positions of the sun and moon the results will be intermediate. Let us now suppose that as the moon revolves the tidal protuberance follows it, then as the earth turns on its axis each place of the surface will have low and high water succeeding each other at intervals of twelve hours. From what has been said it is clear that the highest tides will be at the new and full moon, and the lowest at the times half way between; the former are called *spring-tides*, the latter *neap-tides*. In order to make this theory agree tolerably well with observation, it is necessary to suppose that the axis of the tidal spheroid lags somewhat behind the moon.

ON ECLIPSES.

70. Eclipses are of two kinds, namely, of the moon and of the sun, and may be described in general as being caused respectively, by the passage of the moon through the shadow thrown behind the earth, and by the passage of the moon between the earth and the sun, so as to intercept the sun's light from the earth. It is manifest that if the motions of the sun, earth, and moon, be accurately known so that their relative positions at any time can be predicted, it will be only a matter of calculation to determine when an eclipse will take place; but the method of calculation must be very different in the case of a lunar from that of a solar eclipse, and the latter will be very much more complicated than the former, as will be seen when we have described the phenomena more particularly.

Lunar Eclipse.

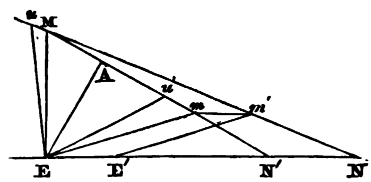
71. Let S be the sun, E the earth, draw the common tangents to their surfaces ACU, BDU, meeting in U, and the two $APDP''$, $BPCP'$, meeting in P between the sun and earth.



Then the portion of the cone, of which the vertex is U, behind the earth is the earth's shadow, and is called the *umbra;* the portion of the cone, of which the vertex is P, behind the earth is partially free from the sun's rays in consequence of the intervention of the earth, and is called the *penumbra.* When the moon is eclipsed it is observed to enter the penumbra first, and afterwards the umbra: on account of the relative magnitude of the earth's diameter and the distance of

the moon, the distance of the vertex of the umbra from the earth is always greater than that of the moon, and hence the moon will always pass through the umbra if at the time of geocentric opposition its path is rightly directed.

72. But there is not necessarily an eclipse of the moon at each opposition, because the moon's orbit is inclined to that of the ecliptic, and at the time of opposition it may not be sufficiently near to a *node*, (or point in which her path crosses the ecliptic,) to pass through the umbra. We shall endeavour to explain the mode of making the calculations to determine whether at any given opposition the moon will be eclipsed, and to what extent.



73. Let EN be a small portion of the path of the centre of the earth's shadow at the distance of the moon, considered as a straight line, MN a portion of the path of the moon's centre, M being the position at the time of opposition, and N the node. Then it will be convenient to suppose the earth to remain fixed, and the moon to move in an imaginary orbit MN', called a *relative* orbit, such that the distance between the centres of the moon and shadow shall always be the same as they actually are. Let m' be the moon's centre when the centre of the shadow is at E'; join $E'm'$, take $m'm$ equal and parallel to EE', and join Em; then Em will be equal and parallel to $E'm'$, and m will therefore be a point in the relative orbit. It is not difficult to see that the relative orbit MN' will be a straight line, and its inclination $MN'E$ to the plane of the ecliptic may be calculated from a knowledge of the inclination of the moon's orbit and of the relative velocities of the earth and moon.

The radius of the umbra at the moon's distance, or rather the angle subtended by that radius at the earth, may be easily calculated from the parallax of the sun and moon and their

apparent diameters, all which quantities are known and registered.

Now draw EA perpendicular to the relative orbit, then since ME, which is the moon's latitude at the time of opposition, and EMA are known, EA may be calculated. EA is the nearest approach of the centres of the moon and the umbra; if then EA be greater than the sum of the radii of the moon and the umbra, there will be no eclipse, if less there will be an eclipse of greater or smaller degree of obscuration according to the value of EA. If we draw two lines Eu, Eu', each equal to the sum of the radii of the moon and umbra, then u and u' will be the positions of the moon's centre at the commencement and termination of the eclipse respectively; and by calculating these positions we can easily determine the time of the commencement and termination.

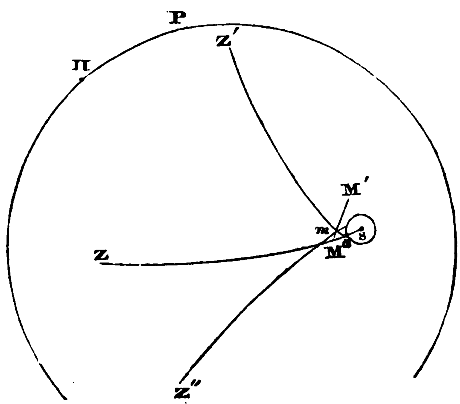
74. Since a lunar eclipse is caused by an actual deprivation of the sun's light, in order to determine the places at which a given lunar eclipse will be visible, it is only necessary to determine the places which will have the moon above their horizon at the time. The calculation is easily made, but for practical purposes it is sufficient to proceed as follows: take a common terrestrial globe, determine upon it the moon's place at the commencement of the eclipse, then all places on the hemisphere lying round this point will see the commencement of the éclipse ; in like manner determine the hemisphere from all places of which the termination is visible; then the whole of the eclipse will be visible from all places which are common to these two hemispheres.

Solar Eclipse.

75. The calculation of a solar eclipse is more difficult than that of a lunar, because the moon by its interposition between the earth and sun will intercept the sun's light from some portions of the earth's surface, but not from others, and the sun may be visible at a given place during an eclipse although the eclipse may not be visible.

We shall attempt to give some account of the mode of calculating the circumstances of a solar eclipse, premising that as we have already supposed the earth to remain fixed and the moon to move in a relative orbit, so here we shall suppose the sun to be fixed and the moon to be apparently depressed by the difference of the solar and lunar parallax, or (as we may call it) the *relative* parallax. This relative parallax, as well as the apparent diameters of the sun and moon are known from the Nautical Almanack, or some equivalent work.

76. Let P be the pole of the equator, and for distinctness' sake, let the plane of the paper be the plane of the solstitial colure: S the sun's centre supposed fixed, MM' the moon's relative orbit; round S draw a small circle, having for its radius the sum of the apparent semidiameters of the sun and moon; then to any place, which is so situated that the moon's centre can be sufficiently depressed by parallax to make it fall within this small circle, the eclipse will be visible.



Let us find the place at which the eclipse will be first seen. Draw SaM an arc of a great circle, and make Ma equal to the relative parallax, and produce SM to Z, so that $Za = 90^0$, then Z will be the zenith of the place required, for to such a place M will be depressed towards S by the whole relative horizontal parallax.

Again, suppose we wish to determine those places on the

earth's surface to which the first contact for different positions of the moon first becomes visible. Let m be any position of the moon in the relative orbit, then from m we can draw, in general, two arcs of great circles mb, mb'*, to meet the small circle round S, and if we produce these to Z' and Z'', making $bZ' = b'Z'' = 90^0$, the moon will appear depressed as much as possible to these two places, and therefore for the position m of the moon Z' Z'' are the zeniths of the two places to which an apparent contact will be first visible. And so we may trace a curve on the earth's surface including all places determined by this construction.

In like manner any other problem may be solved; the most general perhaps is this, to find all places on the earth's surface, at which a given portion of the sun's surface will appear obscured at a given time.

By methods founded upon the principles which we have endeavoured briefly to describe, maps are constructed, such as those in the Nautical Almanack, exhibiting curves on the earth's surface comprehending all places for which an eclipse will be visible in a given degree.

ON THE SATELLITES OF THE PLANETS.

77. The satellites have been already described as secondary bodies, which attend the planets in their orbits, and revolve about them in the same manner as the planets themselves revolve about the sun. The earth is, as we well know, attended by one such satellite, viz. the moon, a body which on account of its peculiar interest to ourselves has already received a separate notice. In these systems Kepler's laws are obeyed, except so far as the motion is disturbed by extraneous causes.

78. The satellites of Jupiter are those which have been most observed; they revolve from west to east, that is, in the same direction as the moon and planets, and in planes nearly coinciding with the equator of the planet. Their eclipses are frequent; in fact, on account of the smallness of the inclinations of their orbits and the great length of Jupiter's

* The letters bb' are omitted in the figure.

shadow, three out of the four are eclipsed at every revolution, usually the fourth also, though on account of the greater inclination of its orbit not invariably.

79. The most remarkable result of observations of Jupiter's satellites is the discovery of the fact of the propagation of light with a finite velocity, and the actual calculation of that velocity. This we proceed to explain.

It was observed by Roemer, that the eclipses of Jupiter's satellites always happened too *soon*, that is, sooner than he expected from an average of observations, when Jupiter was in geocentric *opposition*, and therefore the earth as near to Jupiter as possible, and too *late* when in geocentric *conjunction*, and therefore the earth as far from Jupiter as possible. This was accounted for by the hypothesis that light required a finite time for its propagation, and that therefore the same phenomenon would not appear to happen at the same moment to two observers at stations 190,000,000 miles apart, as is in fact the case with two observers at opposite sides of the earth's orbit. It is easy to see that, supposing the error in the time of the eclipses to be due to this cause, we have sufficient data for calculating the actual velocity of light, and the velocity thus found coincides with that which results from the phenomenon of aberration, which has been already explained (Art. 49): the coincidence of the two independent determinations leaves no doubt respecting the accuracy of each.

ON TIME.

80. There are three kinds of days recognised by astronomers, viz., Sidereal, Solar, and Mean Solar. Each of them is divided into 24 hours, each hour into 60 minutes, and each minute into 60 seconds.

The *sidereal* day is the interval between two successive transits of the true first point of Aries; or it is the time of the earth's revolution on its axis. Sidereal hours are those marked by the sidereal clock already described. (Art. 28.)

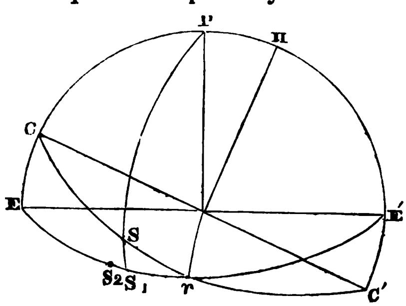
The *solar* day is the interval between two successive transits of the sun, and is therefore not of constant length; hence, although the beginning of this day is marked by a very obvious phenomenon, viz., the transit of the sun, yet it would be extremely inconvenient as the ordinary standard of time.

The *mean solar* day is equal in length to the average of true solar days; its commencement is marked by no actual phenomenon, but by the transit of an imaginary sun which is supposed to move uniformly in the equator with the sun's *mean* or *average* velocity.

Ordinary clocks indicate mean solar time, and are therefore sometimes in advance of the sun, sometimes behind it. The quantity which must be added to, or subtracted from, true solar, to give mean solar time, is called the equation of time, and requires a particular explanation.

81. *The Equation of Time.*

The equation of time may be conceived of as arising from two distinct causes, viz., first, the motion of the sun in the ecliptic, and not in the equator, in consequence of which the time of noon would vary even if the motion of the sun in its orbit were uniform; and, secondly, the variable motion of the sun in its orbit, in consequence of which there would be an equation of time even though the ecliptic had no obliquity. It will give distinctness to our explanation to consider these two parts of the equation separately.



I. Let $E \Upsilon E'$ be the equator, $C \Upsilon C'$ the ecliptic, P, Π their respective poles: and let S be the sun in the ecliptic

between ♈ and the summer solstice. Draw the declination circle PSS_1, then S_1 is the place of the sun in the equator. Now the *mean* sun is to be supposed to start with the true sun from ♈, and to move in the equator with the sun's velocity; hence it is manifest that the two suns will be on the solstitial colure at the same time, and will meet at the autumnal equinox: but between ♈ and the solstice, the R.A. of the mean will be greater than that of the true sun, for the R.A. of the mean sun is to be equal to S♈, which it is easy to see is greater than S_1♈; because, if we take the case of the sun being very near ♈, we may treat S♈S_1 as a plane right-angled triangle, in which the hypothenuse S♈ is the greatest side, and the same will be true until the sun reaches the solstice, when the R.A. of the two is the same. Take then S_2 as the place of the mean sun, S_2 being further from ♈ than S_1. Now the earth revolves about its axis in the same direction as that in which the sun moves, consequently the meridian of any place is brought to S_1 before it reaches S_2, that is, *apparent* noon precedes *mean* noon when the sun is moving from an equinox to a solstice. In like manner it will appear, that the reverse is the case from a solstice to an equinox. Hence the equation of time, so far as it depends upon the obliquity of the ecliptic, is *subtractive* from an equinox to a solstice, *additive* from a solstice to an equinox, and vanishes at both solstice and equinox.

II. In consequence of the eccentricity of the sun's orbit, and the law of Kepler, according to which the areas swept out in equal times are equal, the sun moves with more than its mean velocity when near perigee, and less when near apogee. Suppose a fictitious sun to move in the ecliptic with the sun's mean motion, and to coincide with the real sun in perigee; then from perigee to apogee the real sun will be before the mean sun, and therefore *mean* noon will precede *apparent;* the opposite will be the case when the sun is moving from apogee to perigee: hence from perigee to apogee the equation of time is *additive*, from apogee to perigee *subtractive*, and zero at both apogee and perigee.

To conceive of these two effects as combined, it is only necessary to consider the sun which moves in the ecliptic, and which we spoke of as the real sun when discussing the equation of time arising from the obliquity, to be not the real sun, but a second fictitious sun moving in the ecliptic with the sun's mean motion, and coinciding with the real in perigee and apogee.

82. *To find the Mean Time by observation.*

If we have a fixed observatory, nothing is more simple, because we have only to observe the time of the sun's transit, which will give the time of apparent noon, from which we deduce the time of mean noon by adding or subtracting the equation of time: or we may observe the transit of a known star, which will give us the sidereal time, from which the solar may be deduced; for the sidereal time determines the position of ♈, and since we have tables giving the mean R.A. of the sun, we may hence deduce the position of the mean sun with respect to our meridian at the time of observation.

But when we have no fixed observatory, we must resort to other methods: we shall describe only one. Let an observation of the altitude of the sun be made before noon, and another made when the altitude is the same after noon; then the sun was on the meridian at a period equidistant from the two observations; or, if a clock indicates h and h' hours at the two observations respectively, the time of apparent noon by the clock is $\frac{h + h'}{2}$: apparent noon being known, the mean time may be deduced.

The preceding result is only approximately true, on account of the sun's motion between the two observations; but, if necessary, a correction may be introduced.

83. *On the different kinds of Years.*

Three kinds of years may be reckoned, which we may call astronomical, the *tropical*, the *sidereal*, and the *anoma-*

listic; besides which there is the year in the vulgar acceptation of the word, which we may call the *civil* year.

The *tropical* year is the interval between the two successive arrivals of the sun at the first point of Aries, and its length is about $365^d\ 5^h\ 48^m$.

The *sidereal* year is the interval between the sun's leaving and returning to the same point of the heavens. Its length differs from that of the tropical year on account of the motion of ♈, which has been before explained, and is found to be $365^d\ 0^h\ 9^m$.

The *anomalistic* year is the interval between two successive passages of the sun through perigee, and differs from the preceding on account of a slow motion of the perigee, produced by the disturbance of the planets; its length is $365^d\ 6^h\ 13^m$.

No one of these years would be convenient for ordinary purposes, because a year, to be convenient, should consist of an integral number of days. The *civil* year is therefore made to consist of 365 days, but to prevent the seasons from falling at different periods of the year, every fourth year, or leap-year, is made to consist of 366 days. By this means the average length of a civil year is made to be $365^d\ 6^h$, which is rather more than a tropical year; to correct this error, which in the course of centuries would become considerable, the intercalation is ordered to be omitted in the years completing centuries, when the number of years is not divisible by four; thus only 97 days are intercalated in 400 years instead of 100: the error which remains will not amount to a whole day in 4500 years.

ON THE MODES OF DETERMINING TERRESTRIAL LONGITUDE.

84. The problem of finding the longitude is one of the greatest practical importance, and one which for a long time presented great difficulties. It is evident that if by any means we can ascertain at a given moment what is our own mean time of day, and also what is the mean time at some given place, as Greenwich, we shall know how many degrees

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we are situated to the east or west of Greenwich; now the time of day at any place may be ascertained by observation, as already explained (Art. 82), hence the problem of the longitude reduces itself to that of ascertaining at any given time and place mean Greenwich time.

It is obvious therefore that the problem is solved, if we can obtain watches set to Greenwich time and of sufficient accuracy to be depended upon; and the perfection to which chronometers have been brought, has in fact made the discovery of the longitude at sea a matter of perfect simplicity. Nevertheless there are certain astronomical methods which deserve notice; the following are some of them.

85. *By observation of an eclipse of Jupiter's satellites.*

The time of an eclipse of one of Jupiter's satellites is evidently quite independent of the place from which it is observed; and the motions of the satellites being known, it is possible to predict every eclipse and to register the time of its happening according to Greenwich mean time; if then an observer at another place observes an eclipse and notes the time of its happening, he will be able to compare, by means of the Nautical Almanack, his own mean time with that of Greenwich, and so determine his longitude.

86. *To find the longitude by observing the distance of the moon's centre from certain fixed stars.*

This is a mode which can be practised at sea.

Let the observer with a Hadley's sextant observe the distance of a star from the moon's centre, by noting its distance from the nearest and furthest point of the moon's disk and taking half the sum of these distances. The distance thus found must be corrected for parallax and refraction.

Also, let the observer, immediately after making the preceding observation, (or another observer simultaneously,) take the altitude of the star. From this observation, coupled with the N.P.D. of the star and the latitude, which are supposed

known, the apparent time of the place of observation may be found.

Now the Nautical Almanack gives the distance of the moon from certain stars, of which that observed must be one, for every three hours, and from this we can easily determine the Greenwich time for which the distance of the moon and star is exactly that which we have determined, that is, we can determine the Greenwich time of taking the observation; and the comparison of this with the apparent time at the place of observation determines the longitude.

87. The longitude of a fixed Observatory may be determined thus. Suppose the moon is observed at the Observatory in question to culminate with a certain fixed star, then the same star and the moon will not culminate together to a place on a different meridian, because the moon will have moved in R.A.: suppose that the difference of time (a) between the transit of the moon and the star is observed at the second place, which will be in fact the moon's motion in R.A. Also let A be the motion of the moon in R.A. in the time of a complete revolution of the earth, and x the difference of longitude of the two places, then we shall have

$$x^0 : 360^0 :: a : A;$$

or $x^0 = \frac{a}{A} 360^0$, the difference of longitude required.

ON COMETS.

88. We shall conclude this treatise with a few words on comets. These are in fact planets revolving in conic sections about the sun in their focus, but in orbits of very great eccentricity; the orbits of the greater number are ellipses so much elongated as to be sensibly parabolas, and some seem to move in hyperbolas. The number of these bodies belonging to, or rather infesting, our system, appears to be very great indeed; but few have been sufficiently observed to identify them, when after travelling far into space they again approach

the sun, and become visible to us. Of the few so observed we may notice that which bears the name of Halley's Comet, the period of which is about 75 years; also Encke's Comet, the period of which is $3\frac{1}{2}$ years; and Biela's, $6\frac{3}{4}$ years.

The last two possess remarkable interest; the former from the fact of its period being found to diminish, or its distance from the sun to diminish, a fact which would seem to indicate the existence of some resisting medium in which the planets move, but which has at present not sensibly affected the motions of the larger planets. The latter is remarkable as having been lately discovered to be a double comet, or combination of two comets, probably revolving round the centre of gravity of the two.

The mass of the comets appears to be very small, but they are more easily visible than they otherwise would be from the fact of their throwing out enormous tails of light, which are observed to be always turned away from the sun, but have not as yet been accounted for as physical phenomena.

THE END.